PRACTICAL
PHYSICS

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PRACTICAL PHYSICS

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GALILEO GALILEI. Born 1564, in Pisa, Italy. Died 1642. Often called "the father of modern science" because he was one of the first who thought it worth while to subject his ideas to the test of experiment.

PRACTICAL PHYSICS

FUNDAMENTAL PRINCIPLES AND APPLICATIONS TO DAILY LIFE

BY

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PREFACE

The most difficult problem which confronts any author of a textbook is the selection of material. This is usually a process of exclusion. One has always to keep in mind the capacities and limitations, the interests and inclinations, of the young people most directly concerned, as well as the beauty and vast extent of the subject to be taught. This is especially true of a first course in physics. The number of suitable topics is far greater than can be well handled in any one-year course, however substantial it may be. A good book may, therefore, be judged as well by its *omissions* as in any other way. In preparing this book, we have tried to select only those topics which are of vital interest to young people, whether or not they intend to continue the study of physics in a college course.

In particular, we believe that the chief value of the *informational* side of such a course lies in its applications to the machinery of daily life. Everybody needs to know something about the working of electrical machinery, optical instruments, ships, automobiles, and all those labor-saving devices, such as vacuum cleaners, fireless cookers, pressure cookers, and electric irons, which are found in many modern homes. We have, therefore, drawn as much of our illustrative material as possible from the common devices in modern life. We see no reason why this should detract in the least from the *educational* value of the study of physics, for one can learn to think straight just as well by thinking about an electrical generator, as by thinking about a Geissler tube.

This does not mean that we have tried to make the subject interesting by selecting only the easy topics. There are many parts of physics which are of great practical value, but are essentially difficult. We have tried to present these subjects very slowly and carefully, believing that if any presentation is so simple and direct that the student can understand it clearly, his very understanding begets at once the interest which is fundamental.

Even after a careful exclusion of material, we have selected somewhat more than it is probably advisable for any class to undertake in a single year. This gives the teacher an opportunity to adapt his instruction to the local needs of his community and to the amount of time available. In particular, the chapter on the strength of materials, the discussion of momentum, the chapter on the beginnings of electricity, the chapter on alternating currents, the chapter on electric waves and X-rays, and even the chapter on sound, may well be omitted altogether, or assigned for outside reading without careful discussion, if it seems desirable. We believe that it is most important for teachers to select carefully just what material they can best use, and to teach that *thoroughly*, rather than try to touch upon many topics superficially.

We think it of great importance that the topics in a course in physics should be arranged in the most *teachable order;* that is, with the easiest and simplest topics first. Thus in mechanics, the subject of acceleration and Newton's laws is essentially hard, and so we have put it at the end of that part of the book. On the other hand, the simple machines, such as the lever and the wheel and axle, are essentially easy, and so they come first.

To understand any machine clearly, the student must have clearly in mind the *fundamental principles* involved. Therefore, although we have tried to begin each new topic, however short, with some concrete illustration familiar to young people, we have proceeded, as rapidly as seemed wise,

to a deduction of the general principle. Then, to show how to make use of this principle, we have discussed other practical applications. We have tried to emphasize still further the value of principles, that is, generalizations, in science, by summarizing at the end of each chapter the principles discussed in that chapter. In these summaries we have aimed to make the phrasing brief and vivid so that it may be easily remembered and easily used.

The *problems* are the result of considerable experience in trying to find suitable numerical exercises which will emphasize and illustrate the principles involved, with a minimum of arithmetical drudgery. They, too, are arranged, within each group, as far as possible in the order of their difficulty. It should always be emphasized, however, that the study of physics does not begin and end in the classroom, but is intimately connected with industrial and domestic life. It is very desirable to stimulate in students thought and imagination about what they see, and to get them into the habit of asking intelligent questions of the mechanics, artisans, and engineers whom they meet. We have, therefore, added at the end of each of the earlier chapters, and in many places in the later chapters, *questions*, which require some knowledge gained in this way from outside life. We do not expect that every student can answer even a majority of these questions at first; but after he has tried to answer them, he is in a position to learn a great deal from the subsequent discussion of them in the classroom.

Our treatment of acceleration, Newton's laws, kinetic energy and momentum, is essentially different from either the dyne and poundal method common in physics textbooks, or the "slug" or "wog" method of engineers, and is apparently new. It has, however, been thoroughly tried out in the classroom, and we find it simpler and more direct than the usual presentation. We feel sure that it is as precise and scientific in its logic as any other. It was first developed by

Professor E. V. Huntington, of Harvard University, to whom we gladly make acknowledgment of priority.

We have borrowed ideas also from the books of Mr. Frank M. Gilley, of the Chelsea High School, and Director E. Grimsehl, of Hamburg, Germany. We have received valuable assistance in the preparation of the manuscript from Professor Frank A. Waterman, of Smith College; Mr. Irving O. Palmer, of the Newton Technical High School; Professor J. M. Jameson and Professor J. A. Randall, of Pratt Institute, and many others. To all of these gentlemen we give our hearty thanks.

We are indebted to the General Electric Company, the Westinghouse Companies, the Columbia Graphophone Company, Stone and Webster, and others for material for certain of the plates and illustrations.

Finally, we wish especially to express our obligation to Dr. William C. Collar, lately of the Roxbury Latin School, and to Professor W. S. Franklin, of Lehigh University, but for whose initiative, encouragement, and interest, this book never would have been written.

We shall be grateful for corrections or suggestions from any source.

N. H. B.
H. N. D.

CONTENTS

PRACTICAL PHYSICS

CHAPTER I

INTRODUCTION: WEIGHTS AND MEASURES

Why study physics — content and divisions of physics — physics involves measurement as well as merely description — units in English and metric systems — density.

1. Why study physics? Every one these days has had something to do with machines of one sort or another all his life. In the country we mow, reap, and thresh grain with machines; we pump water with windmills, gas or hot-air engines; and we skim milk with a machine called a separator. In the city we travel on electric cars; we go upstairs on hydraulic or electric elevators; we print our newspapers on presses run by electric motors; and we distribute our mail through pneumatic tubes. In business and in commerce we are constantly using steam, gas, and electric engines, cranes and derricks, locomotives, ships, and automobiles, and perhaps, in a few years, we shall all be using flying machines.

Every one has used some of these devices and almost every one has at some time wondered and perhaps discovered how each of them works. That is, almost every one has already begun to study *physics*, for it is one of the chief aims of physics to discover all that can be known about such machines as have just been mentioned.

2. Physics a science. The sort of physics that will be found in this book differs from the sort that every one has been unconsciously studying all his life, chiefly in that it seeks to answer not only the questions "why" and "how," but also the question "how much." It is only when we

begin to *measure* things definitely that we get the kind of information that helps us to use them to the best advantage. Thus every one knows in a vague way that an automobile goes up a hill because the gasolene which is burned in the engine makes it turn the driving wheels, and these in turn push against the road, if it is not too slippery, and thus propel the automobile. The physicist, when he had thought of all this, would go on to ask himself such questions as "How much gasolene does it take, how much ought it to take under ideal conditions, and what becomes of the difference? How much force must be exerted by the brakes to hold the automobile on a hill, how large a brake surface will do this, and how strong must the brake wire be?" When he can answer all these and many other questions, he is in a position to use his machine more effectively, and perhaps to improve its mechanism.

3. Divisions of physics. The object of studying physics is, then, chiefly to learn to think accurately about very familiar things. But these things are so various in kind that we shall find it convenient to divide the whole subject into five divisions: mechanics, heat, electricity, sound, and light. For example, suppose we wanted to make a thorough study of the automobile. Under **mechanics**, we should study about its cranks, gears, levers, valves, and brakes, including their movements, and the strength of the material of their construction; under **heat**, the engine, its fuel and radiator; under **electricity**, the spark plug, spark coil, magneto, and battery; under **sound**, the horns and trumpets; and finally, under **light**, the lamps and their reflectors and lenses. In a similar way it might be shown that any piece of modern machinery, whether it is an automobile or a locomotive, a motor boat or an Atlantic liner, a flying machine or a submarine boat, is not only an embodiment of the principles of physics, but has in very large measure been made possible by the science of physics.

4. Physics contains some abstractions. While it is true that physics has to do chiefly with familiar things, yet in order to make its study effective we shall also have to consider some things which are not so familiar, such abstractions as density and calories and wave length and refraction and electrical resistance, which may not be interesting at first, and may seem to have little to do with our everyday life. We shall also find many problems to be solved whose answers will seem trivial and unimportant. These things should be done patiently because they pave the way for more valuable things later on.

5. Physics begins with measurements. At the very outset we may well recall an old saying of Plato's: "If arithmetic, mensuration, and weighing be taken away from any art, that which remains will not be much." In the laboratory the student will learn to measure many different kinds of things, not mainly for the sake of the results he gets, but rather that all through life he may know a good measurement when he sees one, and may be able to discuss accurately and with confidence the quantitative problems that are always coming up.

6. Units of measurement. In business in the United States the value of things that are bought and sold is measured in dollars and cents. Fortunately this system of money is made on the decimal plan, that is, in multiples of ten. Our system of weights and measures, on the other hand, is not a decimal system, and is very inconvenient. Nevertheless, since the pound, foot, quart, gallon, and bushel are still in general use in Great Britain and in the United States, we must be familiar with them. During the last century most of the other civilized nations have adopted the **metric** system of weights and measures, in which the relation of the units is expressed in multiples of **ten**. In scientific work the metric system is almost universally used throughout the world, because it greatly reduces the work in making computations.

Therefore it is advisable for us to become proficient in the use of both the English and the metric system of weights and measures.

7. Meter and yard. The **meter** is the distance between two lines on a metal bar (Fig. 1) which is preserved in the vaults of the International Bureau of Weights and Measures near Paris.*

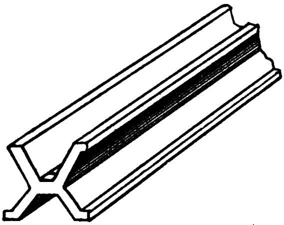


FIG. 1. — The international meter.

Since the length of this metal bar changes a little with the temperature, the distance is measured at the temperature of melting ice. A very accurate copy of this bar is deposited in the United States Bureau of Standards in Washington, D.C., and this copy is the legal meter of the United States.

In the United States the **yard** is legally defined as $\frac{3600}{3937}$ of a meter.

8. Some important units of length. In the problems of physics we shall find that certain units of length are very frequently used. These are given in the following table:

UNITS OF LENGTH

ENGLISH.

1 foot (ft.) = 12 inches (in.).
1 yard (yd.) = 3 feet.
1 mile (mi.) = 5280 feet.

* It was originally intended that the meter should be equal to one ten-millionth part of the distance from the equator to either pole of the earth, but it is impossible to reproduce an accurate copy of the meter on the basis of this definition. Later measurements have shown that the "mean polar quadrant" of the earth is about 10,002,100 meters.

METRIC.

1 meter (m.) = 1000 millimeters (mm.).
1 meter = 100 centimeters (cm.).
1 kilometer (km.) = 1000 meters.
1 inch = 2.540 centimeters.
1 meter = 39.37 inches.

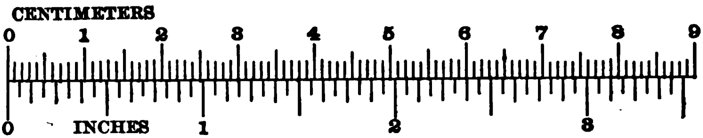


Fig. 2. — Relative sizes of the inch and the centimeter.

9. **Units of area.** The unit of area which is most extensively used is the area of a square of which the side is of unit length. Thus the area of a city house lot is reckoned in square feet, where the unit is a square one foot on each side. In the laboratory, area is often measured in square centimeters (cm^2), the unit being a square one centimeter on each side. It is evident from figure 3 that one square inch is equal to about 6 square centimeters. More accurately, it is 2.54 × 2.54, or 6.45 square centimeters.

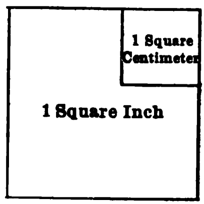


Fig. 3.—Relative sizes of the square inch and the square centimeter.

The usual method of determining area is by calculation from the measured linear dimensions. Thus the area of a rectangle or parallelogram is equal to the base times the altitude ($A = b \times h$). The area of a triangle is equal to ½ the base times the altitude ($A = \frac{1}{2} b \times h$). The area of a circle is equal to 3.14 times the square of the radius ($A = \pi r^2$).

10. **Units of volume or capacity.** The unit of volume that is most extensively used is the volume of a cube of which the edge is of unit length. Thus the volume of a freight car is reckoned in cubic feet, the unit being a cube one foot on each edge. In the laboratory we measure the capacity of a flask in cubic centimeters (cm^3).

UNITS OF VOLUME

ENGLISH.

1 cubic foot (cu. ft.) = 1728 cubic inches (cu. in.).
1 cubic yard (cu. yd.) = 27 cubic feet.
1 gallon (gal.) = 4 quarts (qt.) = 231 cubic inches.

METRIC.

1 liter (l.) = 1000 cubic centimeters (cm^3)
1 cubic meter (m^3) = 1000 liters.
1 liter = 1.06 quarts.

FIG. 4. — A graduated cylinder.

The usual method of determining the volume of a regular solid is by calculation from the measured linear dimensions. Thus to get the volume of a rectangular block of stone, or a box, we find the product, length by width by depth. In the case of a cylindrical figure we compute the area of the circular base (πr^2), and multiply by the height.

For measuring liquids, we ordinarily use a graduated vessel of metal or glass. Thus in the English system we have gallon and quart measures, and for small quantities, fluid ounces (sixteenths of a pint). In the metric system, we have in the laboratory graduated cylinders (Fig. 4) for measuring liquids in cubic centimeters.

PROBLEMS

1. Change 2.55 meters to centimeters.
2. Change 1575 cubic centimeters to liters.
3. A boy is 5 feet 6 inches tall. Express his height in centimeters.
4. Express 1 kilometer as a decimal part of a mile.
5. The Falls of Niagara on the American side are about 165 feet high. Express this in meters.
6. A standard size of automobile tire is 5 inches in diameter and fits a 38-inch wheel. Express these dimensions in the metric system.
7. If you wanted to buy $1\frac{5}{8}$ yards of silk in Paris, what length should you ask for?
8. A certain type of Bleriot monoplane has a wing surface of 15 square meters. Express this in square feet.
9. How many gallons in a cubic foot?
10. Milk sells in Berlin for 40 pfennigs per liter. What is its cost in cents per quart? (100 pfennigs = 1 mark = $0.238.)

11. How many liters does a tank hold which is 3 meters long, 1.5 meters wide, and 1 meter deep?

12. A cylindrical berry box is measured and found to be 6.15 inches in diameter and 2.1 inches deep. What is its capacity in dry quarts? (In the United States it is understood that a dry quart contains 67.2 cubic inches.)

11. Units of weight.* The **kilogram** is the weight of a certain platinum-iridium cylinder that is preserved with the standard meter near Paris, or that of a very accurate copy of this cylinder which is deposited in the United States Bureau of Standards in Washington. It was intended that these cylinders should weigh the same as one liter of pure water, but this has turned out to be not quite true.† It is, however, nearly enough true for our present purposes. Therefore the **gram**, which is the one-thousandth part of a kilogram, is the **weight of one cubic centimeter of water.** It may be helpful to remember that our 5-cent nickel piece weighs 5 grams and our silver half-dollar weighs 12.5 grams.

In the United States the pound avoirdupois is defined legally as $\frac{1}{2.204622}$ of a kilogram.

Units of Weight

ENGLISH.

$$1 \text{ pound (lb.)} = 16 \text{ ounces (oz.).}$$
$$1 \text{ ton (T.)} = 2000 \text{ pounds.}$$

METRIC.

$$1 \text{ gram (g.)} = 1000 \text{ milligrams (mg.).}$$
$$1 \text{ kilogram (kg.)} = 1000 \text{ grams.}$$
$$1 \text{ kilogram} = 2.20 \text{ pounds.}$$

1 cubic foot of water weighs 62.4 pounds.

1 cubic centimeter of water weighs 1 gram.

* The distinction between weight and mass will be made in section 148.

† One liter of pure water at 4° C. weighs 0.999972 kilogram.

12. Weighing machines. The **spring balance** (Fig. 5) is a simple machine for getting the weight of things, or for measuring forces of other kinds, such as the pull exerted by a rope. It consists of a coiled spring, and the force exerted is indicated by the pointer on the scale. The spring balance is very extensively used because of its great convenience, and its indications are close enough for many practical purposes.

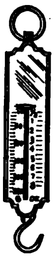


FIG. 5. Spring balance.

The **platform balance** (Fig. 6) consists of a delicately mounted equal-arm balance-beam with pans supported at each end. The balance is used to show the equality of the weights of two bodies; that is, two things are said to have the same weight if they balance each other when supported on the ends of an equal-arm balance. The determination of the weight of any given body by the platform balance depends upon the use of a **set of weights**, which may be combined in such a way as to match the weight of a body.

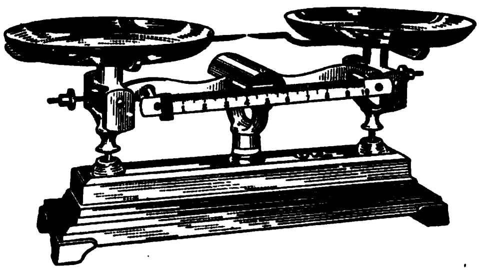


FIG. 6.—Platform balance.

PROBLEMS

1. Change 755 milligrams to grams.
2. Change 1540 grams to kilograms.
3. A girl weighs 52.5 kilograms. Express her weight in pounds.
4. American railways usually allow each passenger 150 pounds baggage. Express this in kilograms.
5. A metric ton is 1000 kilograms. How many pounds is this in excess of the English ton?
6. It is sometimes said, "A pint is a pound, the world around." How much does a pint of water weigh? (1 quart = 2 pints.)

7. A bottle is found to hold 1520 grams of water: (*a*) how many cubic centimeters does it contain? (*b*) how many liters?

8. A boy 5 feet 4 inches tall, and weighing 140 pounds, can walk 3.75 miles in an hour. Express these facts in metric units.

13. Density. Every one knows that lead is "heavier" than cork, and yet the question "which is heavier, a pound of lead or two pounds of cork?" is foolish. The colloquial word "heavy" has two distinct meanings. Two pounds of cork are heavier than one pound of lead in the same sense that two pounds of coal are heavier than one pound of coal. In this case the word "heavy" refers to the total weight of the material. On the other hand, lead is "heavier" than cork in the sense that a piece of lead weighs more than an equal bulk of cork. The word "density" is used to designate more precisely this inherent property of the lead and the cork. That is, lead has a greater density than cork.

The **density** of a substance is its **weight per unit volume.** Thus the density of water is about 62.4 pounds per cubic foot, or 8.34 pounds per gallon. The density of copper is 555 pounds per cubic foot or 0.321 pound per cubic inch. In scientific work it is usual to specify the density of a substance in grams per cubic centimeter (g/cm^3).

Table of Densities

(In grams per cubic centimeter)

Platinum	21.5	Hard woods (seasoned)	0.7–1.1
Gold	19.3	Soft woods (seasoned)	0.4–0.7
Mercury	13.6	Ice	0.911
Lead	11.4	Human body	0.9–1.1
Silver	10.5	Cork	0.25
Copper	8.93	Sulphuric acid (conc.)	1.84
Iron	7.1–7.9	Sea water	1.03
Zinc	7.1	Milk	1.03
Glass	2.4–4.5	Fresh water	1.00
Marble	2.5–2.8	Kerosene	0.8
Granite	2.5–3.0	Gasolene	0.7
Aluminum	2.65	Air	about 0.0012

14. Measurement of density. The simplest way to determine the density of a substance is to weigh the substance and measure its volume.

Thus a piece of pine 6 feet long, 1 foot wide, and 6 inches thick has a volume of 3 cubic feet. If it weighs 90 pounds, its density is 30 pounds per cubic foot.

An empty kerosene can weighs 1.25 pounds, and when filled with kerosene, it weighs 36.25 pounds, so that the net weight of the kerosene in the can is 35 pounds. If the can holds 5 gallons, the density of the kerosene is 7 pounds per gallon.

A block of steel is 15 centimeters long, 6 centimeters wide, and 1.5 centimeters thick and weighs 1050 grams; then the density is $\frac{1050}{135}$ or 7.8 grams per cubic centimeter.

From the preceding examples it will be seen that the density of a body is found by dividing its weight by its volume. In other words,

$$\textbf{Density} = \frac{\textbf{weight}}{\textbf{volume}}.$$

It is also evident that if we know the density of a substance, we can compute the weight of any volume of the substance. It is by this method that engineers calculate the weight of buildings and bridges which it would be impossible to weigh. For example, an engineer finds that a reënforced concrete pier contains 2500 cubic feet of material, and he knows that such material averages 150 pounds per cubic foot. Then the weight of the pier is equal to 2500 times 150, or 375,000 pounds, or about 188 tons. In other words,

$$\textbf{Weight} = \textbf{density} \times \textbf{volume}.$$

If it is the volume of anything that we want to know, we have

$$\textbf{Volume} = \frac{\textbf{weight}}{\textbf{density}}.$$

PROBLEMS

(Use data given in table on page 9 when necessary.)

1. A block of iron is 10 centimeters by 8 centimeters by 5 centimeters, and weighs 3 kilograms. What is its density expressed in grams per cubic centimeter?

2. A block of stone measures 4 feet by 2 feet by 15 inches, and weighs 1625 pounds. Find its density in pounds per cubic foot.

3. How many pounds does 1 cubic foot of aluminum weigh?

4. The cork in a life preserver weighs 20 pounds. What is its volume in cubic feet?

5. A flask with a capacity of 120 cubic centimeters is filled with mercury. How many kilograms of mercury does it hold?

6. A quart bottle is weighed empty and then full of milk. How many pounds should it gain in weight?

7. A cylindrical railway water tank measures on the inside 10 feet in depth and 6 feet in diameter. How many tons of water does it hold?

8. A piece of platinum wire is 12.5 centimeters long and 0.8 millimeter in diameter. How much would it cost if the price of platinum is $1.00 per gram?

9. If a certain copper telephone wire is 0.165 inch in diameter, what does a mile of the wire weigh?

10. The inside diameter of a lead pipe is 1 inch, and the wall is 0.25 inch thick. How many pounds does it weigh per foot?

15. The three fundamental units. On account of its more convenient size, the **centimeter**, instead of the meter, is universally used in scientific work as the fundamental unit of length. For a similar reason, the **gram**, instead of the kilogram, is used as the fundamental unit of weight. The **second** is taken among all civilized nations as the standard unit of time. It is $\frac{1}{86400}$ of the time from noon to noon.

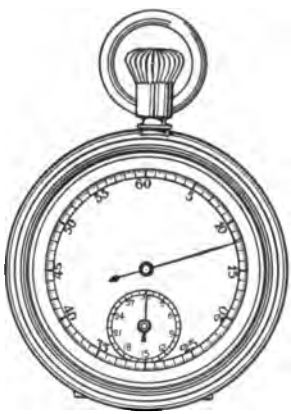


FIG. 7. — Stop watch.

The process of weighing something on a balance is quite dis-

tinct from the measurement of a length, and the measurement of time is wholly different from the measurement either of length or weight. Moreover, each is done with a distinct sort of instrument. In a time measurement the instrument is a clock or watch. For short intervals of time a special type of watch is used, known as a stop watch (Fig. 7).

It is found that the measurement of any quantity, such as the steam pressure in a boiler, the speed of an express train, or the loudness of a foghorn, can, in the ultimate analysis, be reduced to measurements of length, weight, and time. The units of length, weight, and time are therefore the three **fundamental units** of physics.

SUMMARY OF PRINCIPLES IN CHAPTER I

$$\textbf{Density} = \frac{\textbf{weight}}{\textbf{volume}}.$$

QUESTIONS *

1. What is the origin of the prefixes *deci*, *centi*, and *milli*, used in the metric system?

2. How could you determine the volume of an irregular piece of rock by means of a graduated cylinder partly filled with water?

3. How would you measure the diameter of a steel ball?

4. How does your local jeweler get "standard time" to set his clocks and watches correctly?

5. How can the thickness of this sheet of paper be measured?

6. What is the difference between a ship's chronometer and a dollar alarm clock?

7. Why is it that the United States and Great Britain are the only two civilized countries that do not use the metric system commercially?

* In trying to find the answers to these questions, the student is expected to consult various reference books, such as dictionaries, encyclopedias, engineering handbooks, and popular science magazines. He is also expected to keep his eyes open outside of the classroom, and to ask questions of mechanics and tradespeople.

CHAPTER II

SIMPLE MACHINES

Levers of various kinds — principle of moments — force at the fulcrum — weight of a lever — center of gravity in general — wheel and axle — pulley systems — parallel forces.

Work — principle of work — differential pulley — inclined plane — wedge — screw — combinations of simple machines — power — transmission of power.

Friction — so-called "laws of friction" — coefficient of friction — advantages of friction — rolling friction — efficiency of machines.

16. Why we use machines. A man can lift a piano up to a window on the second floor with a rope and tackle. A boy can roll a barrel of flour up into a wagon with a skid. A girl can pull a nail out of a box with a claw hammer although she could not move the nail at all with her fingers alone. It is obvious that we can do many things with simple machines that it would be quite impossible for us to do without them because we are not strong enough. Furthermore, some machines enable us to do things more quickly or more conveniently than we could without them. Most important of all, we often use machines in order to make use of forces exerted by animals, wind, water, or steam.

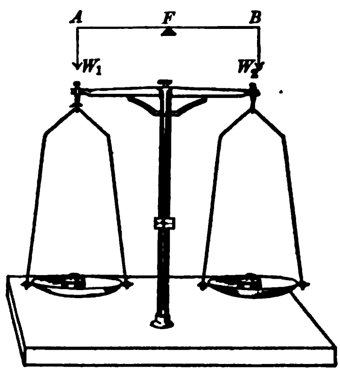


FIG. 8. — Equal-arm lever.

17. Equal-arm lever. — Doubtless the simplest machine is the lever with equal

arms, such as a seesaw, or the walking beam on a steamboat, or the scale beam on a platform balance. In this case we know that equal weights or equal forces just balance when placed at equal distances from the point of support. Thus in figure 8, when W_1 equals W_2, the distance AF must equal the distance BF. In the technical language of physics the point of support (F) of a lever is called its **fulcrum**.

18. Unequal-arm lever. Very often the distances of the weights from the fulcrum are not equal. For example, the distances are unequal when two persons of unequal weight are seesawing, or in the case of an ordinary pump handle. It is evident that at equal distances, the larger weight would have the greater tendency to tip the lever, and also that, with equal weights, the weight at the greater distance from the fulcrum has the greater tendency to tip the lever. Therefore in order to have two unequal weights balance, they must be so placed that the smaller weight is at the greater distance from the fulcrum.

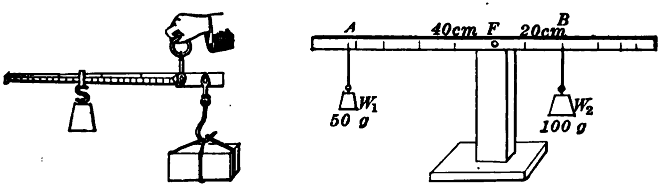


FIG. 9.—Two unequal forces.

If we balance an ordinary meter stick in the middle and suspend a 50-gram weight (W_1) at A, which is 40 centimeters from the fulcrum (F), and then hang a 100-gram weight (W_2) on the other side at such a point as just to balance the first weight, we shall find that the point B where the 100-gram weight is hung is about 20 centimeters from F or half as far from the fulcrum as the 50-gram weight.

Careful experiments show that any two unequal forces will balance only if *the force on one side multiplied by the perpendicular distance of its line of action from the fulcrum equals*

the force on the other side multiplied by the distance of its line of action from the fulcrum. For example, in figure 9,

$$W_1 \times AF = W_2 \times BF.$$

This relation of the forces and distances may also be expressed by a proportion

$$W_1 : W_2 :: BF : AF,$$

which may be stated in words as follows: the forces are *inversely* proportional to their distances from the fulcrum. This means that if one force is *three* times as great as another, then its line of action must be *one third* as far away from the fulcrum as the other to make the lever balance.

Crowbars, shears, glove stretchers, pliers, etc., are all examples of this sort of lever.

19. One-arm lever. When the fulcrum is located at one end of the lever, as in figure 10, the same principle is involved. There are two tendencies which must balance, the tendency of the weight to tip the lever down and the tendency of the pull applied to the lever to lift it up. The weight multiplied by the perpendicular distance from the fulcrum to its line of action measures its turning effect about the fulcrum; that is, its tendency to tip the lever down. This must be balanced by an equal turning effect in the opposite direction, namely, the upward pull multiplied by its distance from the fulcrum.

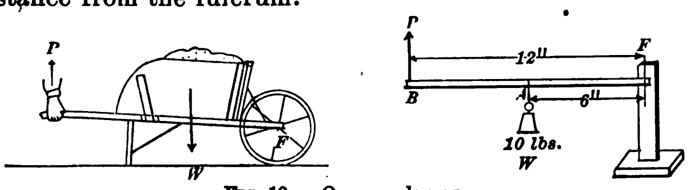


FIG. 10. — One-arm levers.

Suppose we fasten a stick (Fig. 10) by a screw (F) to an upright support, so that the stick is free to turn, and hang a weight (W), say 10

pounds, at a distance of 6 inches from the fulcrum (F) Then if we pull up with a spring balance at a point (b) 12 inches from the fulcrum (F), we shall find that the pull measured by the spring balance is about 5 pounds. (Of course allowance has to be made for the weight of the stick.)

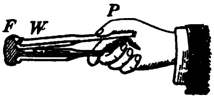


FIG. 11. — Nut cracker.

The equation representing these tendencies to turn the stick in opposite directions would be as before

$$W \times AF = P \times BF.$$

It is also evident that if the 10-pound weight (W) were hung 12 inches from the fulcrum (F), and the upward pull applied 6 inches from the fulcrum, the pull needed would be 20 pounds. In other words, the same principle applies to the one-arm lever wherever the weight and the upward pull are applied.

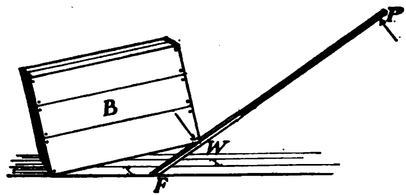


FIG. 12. — Crowbar.

A nut cracker (Fig. 11), a crow bar when used with one end on the ground (Fig. 12), and the forearm (Fig. 13) when it supports a weight in the extended hand are examples of levers with the fulcrum at one end, or one-arm levers.

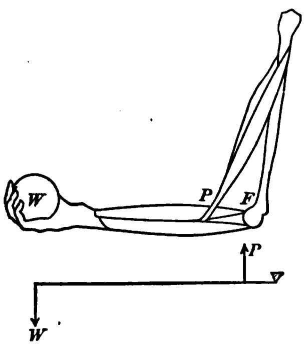


FIG. 13. — Forearm.

20. Lever with two weights. The ordinary wheelbarrow is a good example of a one-arm lever. The fulcrum is located at the axle of the wheel, the weight is the load carried, and the upward pull is exerted on the handles by the

man. In practice, however, it often happens that the load consists of two weights, such as two bags of cement or two boxes or kegs, as shown in figure 14. To get the upward pull we have merely to compute the turning effect of each of the weights (W_1 and W_2) about the fulcrum (F) and make the sum of these effects equal to the turning effect of the upward pull (P). That is,

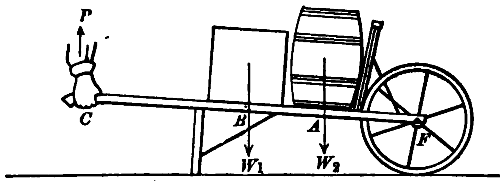


FIG. 14. — Wheelbarrow with two weights.

$$W_1 \times BF + W_2 \times AF = P \times CF.$$

In general, then, we see that we can balance the turning effect of two or more weights by multiplying each weight by the perpendicular distance of its line of action from the fulcrum, and making the sum of these products equal to the product of the pull by the perpendicular distance of its line of action from the fulcrum.

21. Principle of moments. It has been seen that the turning effect of a force depends on two factors, the *amount* of the *force* and the *distance* of its line of action from the fulcrum. This product — **force times perpendicular distance to fulcrum** — is called the **moment** of the force. For a lever to be in equilibrium, *the sum of the moments of the forces tending to turn it in one direction must equal the sum of the moments of the forces tending to turn it in the opposite direction.*

22. Force at the fulcrum. It must not be forgotten that in examples of the lever, the fulcrum itself exerts a force, that is, a push or a pull. When the fulcrum is between the two weights, it evidently has to push up an amount equal to the sum of the weights; that is (Fig. 15, top), $F = W_1 + W_2$.

When the fulcrum is at one end of the lever, and the pull at the other end, it is clear that the fulcrum must exert an upward push, which must be such that it and the upward pull (P) are together equal to the weight. That is (Fig. 15, middle), $W = F + P$. When the fulcrum is at one end and the weight at the other end, it will be readily seen that the fulcrum has to push downward and that the upward pull (P) must equal the sum of this downward push at F and the weight W. That is (Fig. 15, bottom), $P = W + F$. In short, it will be seen that in all these cases *the sum of the forces pulling up must equal the sum of the forces pulling down.*

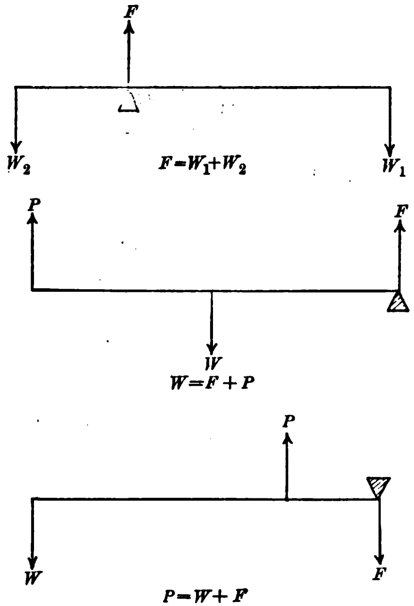


FIG. 15. — Force exerted by fulcrum of lever.

PROBLEMS

1. Identify the fulcrum, and the direction of the two forces, in the case of a pair of shears, a glove stretcher, a pair of tongs, and a nut cracker, regarded as examples of the lever, and think of other examples.

2. What weight placed 20 inches from the fulcrum will balance 100 pounds placed 8 inches away on the opposite side? What is the pressure on the fulcrum?

3. In figure 9 the movable weight on an old-fashioned steelyard weighs 3 pounds, and is placed at such a distance as to balance a 50-pound sack which is hung from a point 1 inch from the point of support. How far from the point of support must the sliding weight be placed?

4. A piece of wire, which is to be cut with shears, is placed 0.5 inch from the rivet. If a force of 25 pounds is applied on the handles 6 inches from the rivet, how much force is exerted on the wire?

5. A plank 12 feet long is to be used as a seesaw by two boys who weigh 100 pounds and 140 pounds. How far from the lighter boy must the prop be placed?

(HINT. — Let x = distance from small boy and $12 - x$ = distance from big boy.)

6. The handles of a wheelbarrow (Fig. 10) are 4 feet 6 inches from the axle, and the load of 200 pounds can be considered as 18 inches from the axle. How much effort must be exerted to raise the handles?

23. Bent lever. Consider next a claw hammer (Fig. 16) with a 12-inch handle. If a 60-pound pull (P) at B is necessary to pull the nail at A, which is 1.5 inches from F, what is the resistance (R) which the nail offers? The moment of P is $P \times BF$, and the moment of R is $R \times AF$, therefore $60 \times 12 = R \times 1.5$, and $R = 480$ pounds. In this case it will be seen that the two arms of the lever are inclined to each other, but the principle of moments applies just as if it were a straight lever. The bent lever is very common as a *part* of a machine.

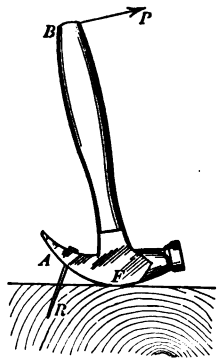


FIG. 16. — Hammer as a bent lever.

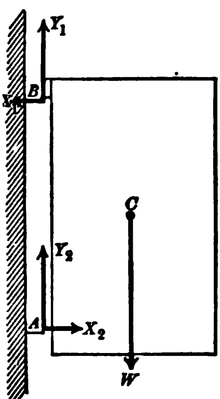


FIG. 17.— Door as a bent lever.

A great many other objects can be regarded as bent levers. For example, suppose a door (Fig. 17) 8 feet high and 4 feet wide, weighing 60 pounds, swings on hinges placed 1 foot from the top and 1 foot from the bottom. (a) What is the vertical pressure on each hinge? If the door is properly hung, the weight will be equally divided between the two hinges, and each hinge will support 30 pounds. (b) How great is

the horizontal *pull* on the upper hinge? If we consider the entire weight of the door as acting at its center, then the moment of this weight about the lower hinge will be 2 times 60, or 120. The moment of the pull exerted by the upper hinge, reckoned about the lower hinge, will be 6 times the pull. Making these two moments equal, we find that the pull is 20 pounds. (*c*) In a similar way, by considering the upper hinge as a fulcrum, we may compute the *push* exerted by the lower hinge, which also equals 20 pounds.

24. Center of gravity. So far in our study of levers we have assumed that the weight of the lever itself could be neglected, but in practice this is not always the case. It is our problem now to find how to make allowance for the weight of the lever.

We have already seen that a lever carrying two weights (Fig. 18) can be supported at a point in between, which we have called the fulcrum, but which we may now call the "center of gravity" or "center of weight." The force necessary to support this point is the same as if the whole weight were concentrated there. In the same way we could support a bar carrying three or more weights on a single fulcrum, if it is placed at the right point. That point would be the center of gravity of the weights. In general, everything has a center of gravity at which we can consider its whole weight concentrated. To find the position of the center of gravity, we have simply to find the point at which the object would balance on a knife edge. This may be computed, but it is usually easier to locate it experimentally.

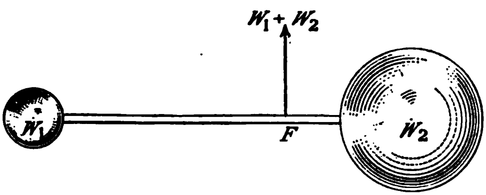


FIG. 18. — Center of gravity of two weights.

25. How to find a center of gravity by experiment. If the shape of the object is simple and its density is everywhere the same, as in the case of a shaft or a board, we should ex-

pect the center of gravity to be in the middle, and if we try to balance the object on some sharp edge, we find that the center of gravity is indeed located at the geometrical center. In the case of an irregularly shaped object like a baseball bat, the simplest way is to balance the bat on a knife edge. In the case of a chair, the center of gravity may be found by considering that, if the chair is hung so as to swing freely, the center of gravity will lie directly under the point of suspension. Therefore, if a chair, or any irregular object, is hung from two points successively, the point of intersection of the plumb lines from these points will locate the center of gravity.

To make this clear, let us take an irregular sheet of zinc and drill three holes near the edge, *A*, *B*, and *C*, in figure 19. Let the zinc be hung from a pin put through the hole *A* and let a plumb line be also hung from the pin. Draw a line on the zinc to show where the plumb line crosses it. Then let the zinc be hung from another hole and draw another line in a similar way. The point of intersection is the center of gravity. When the zinc is hung from the third hole, the plumb line will pass through the center of gravity already found.

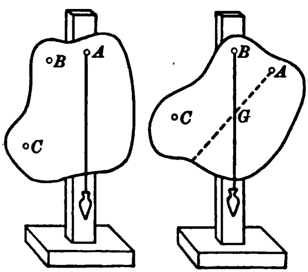


FIG. 19.—Finding center of gravity.

In the case of a ring, or a cup, or a boat, the center of gravity will not lie in the substance itself, but in the empty space inside; but this will not bother us in answering questions about how such objects act. We may, if we like, think of such a center of gravity as rigidly attached to the object by a very light, stiff framework.

We shall find this idea of the **center of gravity** especially convenient in problems where the weight of a lever has to be considered, for we can now assume that the *whole weight of the lever is concentrated and acting at its center of gravity.*

Suppose that an 18-ounce hammer balances 10 inches from the handle end. When a fish is tied to the end of the handle, the whole balances 6 inches from the end. How much does the fish weigh? We may consider the weight of the hammer, 18 ounces, as concentrated at a point 10 inches from the end of the handle or 4 inches from the fulcrum. Let x be the weight of the fish, which is applied 6 inches from the fulcrum. Then we have

$$6x = 4 \times 18,$$
$$x = 12 \text{ ounces, the weight of the fish.}$$

Problems

1. A boy has a 2-pound fish pole 10 feet long, the center of gravity of which is 3.5 feet from the thick end. He finds the weight of his string of fish by hanging them from the thick end of the pole and then balancing the pole on a fence rail. He finds that it balances at a point 15 inches from the end. How many pounds of fish has he?

2. A pole 20 feet long weighs 120 pounds. When a 30-pound bag of meal is hung at one end, the balancing point is 3 feet from the same end. Where is the center of gravity of the pole?

3. A 6-foot crowbar balances at a point 2.5 feet from its sharp end. If a weight of 30 pounds is hung 0.5 feet from this end, and 50 pounds is hung 1 foot from the other end, it balances at its mid-point. How heavy is the bar?

4. A uniform beam AB, 20 feet long, weighing 600 pounds, is supported by props placed under its ends. Four feet from prop A, a weight of 200 pounds is suspended. Find the pressure on each prop.

(Hint. — Regard as a lever, with its fulcrum at one end.)

5. A rectangular gate 3.5 feet high and 5 feet wide has its center of gravity at its geometrical center. It is hung on hinges placed 3 inches from the top and bottom. The gate weighs 100 pounds. (*a*) What vertical pressure should each hinge sustain? (*b*) What is the horizontal *pull* on the upper hinge? (*c*) What is the horizontal *push* against the lower hinge?

26. Wheel and axle. A special form of lever consists of a wheel or crank which is fastened rigidly to an axle or drum. The weight to be lifted, or the resisting force of whatever kind, is generally applied to the axle by means of a rope or chain, and the "effort," or pull, is exerted on the rim of the

wheel, as shown in figure 20. In calculating the effort (P) needed to balance a given resistance (W) we have merely to take moments about the center (F) of the wheel and axle. If we call the radius of the wheel R and that of the axle r, then,

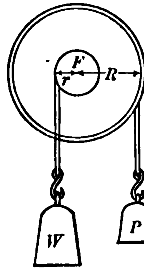


FIG. 20. — Wheel and axle.

Weight × axle-radius = effort × wheel-radius

or $$W \times r = P \times R,$$

or $$\frac{W}{P} = \frac{R}{r}.$$

27. Uses of the wheel and axle. A **windlass** used in drawing water from a well (Fig. 21) by means of a rope and bucket is an application of the principle of the wheel and axle. In the windlass, a crank takes the place of a wheel, and the length of the crank is the radius of the wheel.

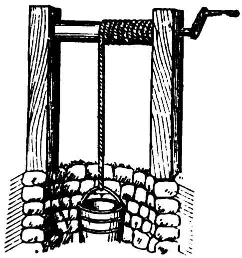


FIG. 21. — Windlass for a well.

If a **wheel** is used in turning the rudder of a boat, the rope attached to the rudder is wound round the axle, and the steersman applies his effort to the handles which project from the rim of the wheel (Fig. 22).

In the **derrick** (Fig. 23), which is used in lifting heavy

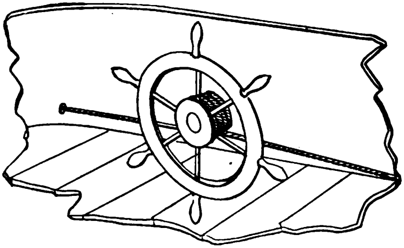


FIG. 22. — Steering wheel in a boat.

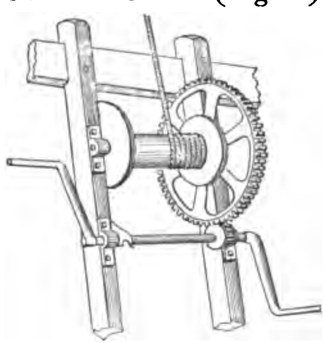


FIG. 23. — Hoisting derrick.

weights, we usually have a double wheel and axle. The effort of the workmen is applied at the cranks, which are attached to one axle. This then drives, through a spur gear, a wheel on a second axle.

28. The pulley. The **fixed pulley**, shown in figure 24, consists of a wheel with a grooved rim, called a **sheave**, free to rotate on an axle which is supported in a fixed **block**. A flexible rope or cable passes over the wheel. It is evident

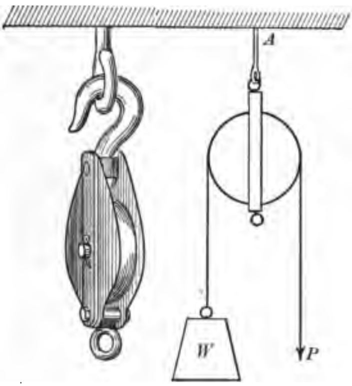


FIG. 24. — Fixed pulley.

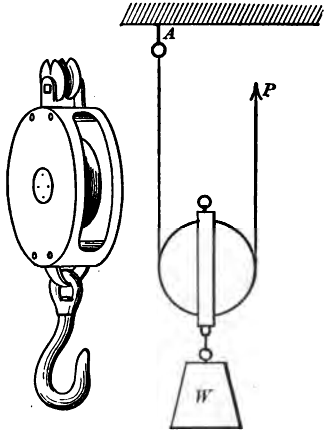


FIG. 25. — Movable pulley.

that if equal weights or equal forces are applied to the ends of the rope, they just balance each other. That is, the effort P is equal to the resistance W. So there is no advantage in the fixed pulley, except that it is sometimes more convenient to exert a certain pull downwards rather than upwards.

Oftentimes the block is attached to the weight to be lifted, as shown in figure 25, and then it is called a **movable pulley**. Here the effort P is not equal to the weight W, for it will be seen that the load W is supported by *two* ropes, and therefore each exerts a pull equal to *one half* the weight. That is,

$$P = \frac{1}{2} W, \text{ or } W/P = 2.$$

The ratio of the weight or resistance to be overcome to the effort put forth is called the **mechanical advantage** *of a machine.* For example, the mechanical advantage of a single fixed pulley is 1 and of a single movable pulley is 2.

29. Combinations of pulleys. In practical work it is quite common to use a fixed block with two sheaves and a movable block with two sheaves, as shown in figure 26. One end of the rope is attached to the fixed block, and the effort is applied to the other end of the rope. Let us compute the relation between the weight to be lifted and the effort applied. From figure 26 it will be seen that the weight and the movable block are supported by *four* ropes, and so the pull on each rope, neglecting the weight of the block, is *one fourth* the weight W. It will also be seen that the pull P is equal to that in each of the ropes, since a pulley only changes the direction of the pull. Therefore

$$P = \tfrac{1}{4}\, W,$$

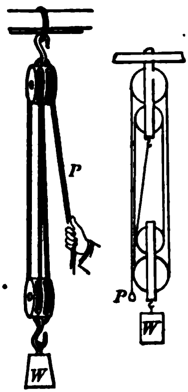


FIG. 26. — Double blocks.

and the mechanical advantage, W/P, is 4.

This means that, neglecting friction and the weight of the movable block, a pull of 100 pounds applied at P would just balance a weight of 400 pounds at W.

In general, we can find the mechanical advantage of any combination of pulleys by counting the number of ropes which support the weight.

30. Parallel forces. Suppose we have a 3000-pound automobile standing on a bridge in a position one fourth of the length of the bridge from one end (Fig. 28), and we wish to know how much of the weight is borne by the supports at each end of the bridge.

First let us try a very simple experiment which will make clear the principles involved in this problem.

Hang a light stick (Fig. 27) by two or more stirrups attached to spring balances (A, B, C), and let several weights (D, E) be hung from it at various points. If the supports do not break, the stick will remain suspended motionless indefinitely. The sum of the forces pulling up is equal to the sum of the forces pulling down. Now suppose that there happen to be several holes through the stick and that a nail is carefully driven through one of them into the wall behind. If the stick did not move before, it certainly will not move now. But we can now think of the stick as a lever with the nail as a fulcrum, and it is in equilibrium about that nail. This means that the sum of the moments of the forces tending to turn it in one direction equals the sum of the moments of the forces tending to turn it in the opposite direction.

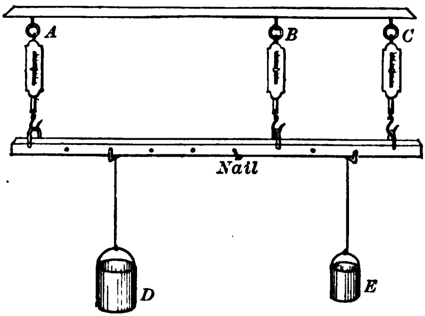


FIG. 27. — Parallel forces.

Evidently this nail could have been put through a hole at *any* point along the stick, and the moments calculated around that point would balance.

This example and section 22 show that when several parallel forces are in equilibrium *two* conditions must be fulfilled.

(1) *The sum of the forces pulling in one direction must equal the sum of those pulling in the opposite direction.*

(2) *The sum of the moments tending to rotate the whole in one direction around any point whatever must equal the sum of the moments tending to rotate the whole in the opposite direction around that same point.*

Let us apply these principles of parallel forces to the problem of the automobile standing on the bridge. This can be represented by figure 28, where A is the weight of the

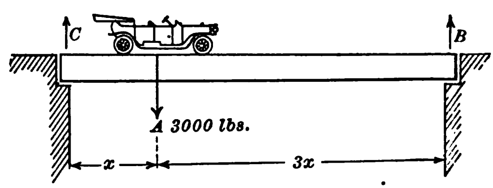


FIG. 28. — Bridge with automobile on it.

automobile, and B and C are the upward forces exerted by the end supports. We know one force ($A = 3000$ pounds), and the relative distances between these forces. We are to find the magnitudes of B and C. Since $B + C = 3000$, $C = 3000 - B$. Suppose we take the position of the automobile as the point about which to compute moments; then we have

$$B \times 3\,x = (3000 - B) \times x,$$
$$B = 750 \text{ pounds, } B\text{'s load,}$$
$$3000 - B = 2250 \text{ pounds, } C\text{'s load.}$$

We can also solve this problem by taking moments first around one end, and then around the other. Working in this way, we do not need the first principle at all. Do this and see if you get the same answers.

It should be noticed that all the machines so far considered, namely, the lever (except the bent lever), the wheel and axle, and the pulley, are simply special cases of parallel forces, and that we can discover anything we want to know about any of them, by means of one or both of the general principles mentioned just above. For the lever and the wheel and axle, the principle of moments is enough, unless we want to know the force at the fulcrum. For that we need the first principle. For the pulley we need only the first principle.

PROBLEMS

1. The diameter of an axle is 1 foot, and the diameter of the circle in which a crank on the axle moves, is 3 feet. If 150 pounds is the weight to be raised, how much force must be applied to the crank?

2. The crank on a grindstone is 9 inches long, and the diameter of the stone is 30 inches. If 50 pounds is the force applied on the crank, what force can be exerted on the rim of the stone?

3. What must be the ratio of the diameters of a wheel and axle, in order that 150 pounds may support 1 ton? What is the mechanical advantage?

4. Two single fixed pulleys are used to raise a barrel of flour, as shown in figure 29. If a barrel of flour weighs 200 pounds, how much does the horse have to pull?

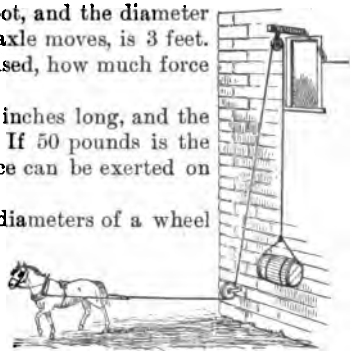


FIG. 29. — Simple pulley system.

5. The gaff of a boat is to be raised by means of a movable single block attached to it, and a fixed double block attached to the top of the mast, one end of the rope being tied to the movable block. How much resistance can be overcome by 100 pounds exerted on the rope?

6. A pair of triple blocks contain three sheaves each. The rope is attached to the upper fixed block. What force is just sufficient to balance a weight of 1 ton, neglecting friction?

7. An automobile gets stuck in the sand. In order to pull it out, a horse, a rope, and a couple of triple blocks are used. If the horse exerts a steady pull of 500 pounds on the rope, and one block is fastened to a tree and the other to the machine, how much resistance can be overcome? Find two solutions for this problem, the rope being fastened in one case to the fixed block, and in the other to the movable block.

8. Two boys, *A* and *B*, are carrying a 100-pound load slung on a pole between them. Their hands are 10 feet apart, and the load is 3 feet from *A*. How much does each carry? Neglect the weight of the pole.

9. A man holds a shovelful of coal, weighing 50 pounds, with his left hand at the end of the shovel, and his right hand 22 inches away. Supposing the center of gravity of the shovel and coal to be 40 inches from his left hand, how much does he push down with his left hand, and how much does he pull up with his right hand?

10. A man and a boy carry a load of 200 pounds on a pole 8 feet long. Where must the load be placed if the boy is to bear only 45 pounds of it?

31. Work. The function of every machine is to do a certain amount of **work.** Now in the technical language of science, **work means the overcoming of resistance.** For example, we do work when we lift a box from the floor to the table, or when we push the box along the floor against friction. But we are not doing work in the scientific sense of the word, no matter how hard we push or pull, if we do not lift or move the box. In other words, work is measured by accomplishment, not by effort or by fatigue.

If we lift one pound one foot, we are said to do **one foot pound** of work; if we lift 5 pounds 3 feet, we do 15 foot pounds of work; or if we pull hard enough on a box to lift

5 pounds and thus drag it 3 feet, we still do 15 foot pounds of work. In other words,

Work (foot pounds) = force (pounds) × distance (feet).

It should be remembered that the distance must be measured in the *same direction* as that in which the force is exerted. Thus, if a machinist exerts upon a file a force of 10 pounds downward and 15 pounds forward, how much work will he do in 40 horizontal strokes, each 6 inches long? Evidently the total distance is 20 feet and the *horizontal* force is 15 pounds; therefore the work done is 300 foot pounds. The vertical pressure does not enter into the calculation of work because the motion is horizontal.

32. Principle of work. In every machine a certain resistance is overcome by a certain effort exerted on another part of the machine. The principle of work which applies to all machines where the losses due to friction may be neglected, may be stated as follows: *The work put into a machine is equal to the work got out.* In short,

Input = output.

For example, in the wheel and axle (see Fig. 20, section 26) the *output* is equal to the weight times the distance it is lifted, and the *input* is equal to the effort times the distance through which it is exerted. For convenience, suppose the wheel makes just one turn. Then the distance the weight is lifted is equal to the circumference of the axle, $2\pi r$, and the distance through which the effort is exerted, is the circumference of the wheel, $2\pi R$. The *input* is $P \times 2\pi R$, and the *output* is $W \times 2\pi r$. Therefore, by the principle of work,

$$P \times 2\pi R = W \times 2\pi r,$$

or

$$P \times R = W \times r,$$

which is exactly the equation got by considering the wheel and axle as a modified lever.

Another example is the system of pulleys shown in figure 26 in section 29. The *output* is equal to the weight W times the distance it is lifted, and the *input* is equal to the effort P times the distance through which it is exerted. Suppose the distance the weight W is lifted is D,

and the distance through which the effort P is exerted is d. The *output* is $W \times D$ and the *input* is $P \times d$. Then, by the principle of work,

$$W \times D = P \times d,$$

or

$$\frac{W}{P} = \frac{d}{D}.$$

But when the weight is lifted 1 foot, it is evident that each of the supporting ropes must be shortened by 1 foot, and therefore P must move 4 feet; in other words,

$$d = 4D.$$

Substituting this value of d in the preceding equation, we have

$$\frac{W}{P} = 4,$$

which is the same as the result which we got by considering the pulley as a case of parallel forces.

33. The differential pulley. In shops where heavy machinery is to be lifted, constant use is made of the **differential pulley,** shown in figure 30. This consists of two sheaves of different diameters in the upper block rigidly fastened together, and one sheave in the lower block. An endless chain runs over these blocks. The rims of the sheaves have projections which fit between the links and so keep the chain from slipping. Such a differential pulley has a *very large mechanical advantage*.

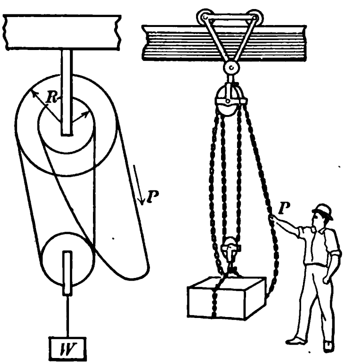


FIG. 30.—Differential pulley.

To see just now it comes to have a large mechanical advantage, let us set up such a pulley and study it carefully. When the chain is pulled down as shown in the diagram, it is wound up faster on the large fixed pulley than it is unwound on the smaller pulley. In order to compute the mechanical advantage of the contrivance, let us suppose that P moves

down far enough to turn the fixed pulley around once. If R is the radius of the large fixed pulley, then the work done by P will be $P \times 2\pi R$. If r is the radius of the small fixed pulley, then the length of chain unwound in one revolution will be $2\pi r$. The weight W will therefore be raised $\frac{1}{2}(2\pi R - 2\pi r)$ or $\pi(R - r)$ and the work done will be $W \times \pi(R - r)$. Therefore, if we neglect losses due to friction, we have

$$W \times \pi(R - r) = P \times 2\pi R,$$

whence,

$$\frac{W}{P} = \frac{2R}{R - r}.$$

Since the difference between the radii of the two fixed pulleys $(R - r)$ is small, it is evident that the mechanical advantage is large.

The differential pulley has a second practical advantage in that there is always enough friction to keep the weight from dropping when the force P is released.

PROBLEMS

1. A man carries in baskets a ton of coal up 20 steps, each 7 inches high. How much work does he do on the coal?

2. In the metric system, work is measured in kilogram meters. How much work is done in pumping 50 liters of water 40 meters high?

3. A man weighing 150 pounds raises himself up a mast in a sling by means of a rope passing over a fixed pulley attached to the top of the mast. If the mast is 100 feet high, how much work does he do? How hard must he pull?

4. If in problem 1 on page 27 the weight is raised 10 feet, how many foot pounds of work are done by the machine?

5. If in problem 2 on page 27 the stone is turned 30 revolutions per minute, how many foot pounds of work are put into the grindstone per minute?

6. In a certain differential pulley the large wheel is 6 inches and the small wheel 5 inches in diameter. What is the mechanical advantage?

34. Inclined plane. Barrels and casks which are too heavy to lift from the ground into a wagon are often rolled up a plank or skid. This is an example of what is called an inclined plane. Every street or road which is not level is an example of an inclined plane. Experience teaches us that

the steeper the incline, the greater the pull required to haul the load up the grade. In order to find out just how the effort and the weight or load are related to the grade, let us try a simple experiment, where friction can be neglected.

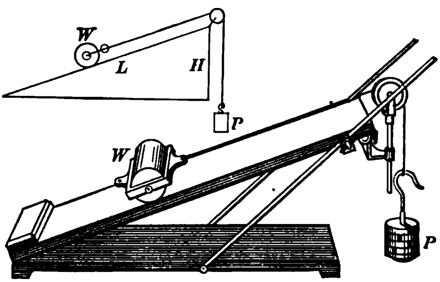


FIG. 31. — Inclined plane.

Suppose we arrange a very smooth plane, such as a piece of plate glass, at an angle, as shown in figure 31. Let the weight or load (W) be a heavy metal cylinder which turns with very little friction. Attach to the cylinder a cord and pass it over a good pulley fastened to the top of the plane, and then hang from the other end enough weights to pull the load slowly up the inclined plane. You will find that the ratio P/W is approximately the same as the ratio H/L, where H is the height of the incline and L is its length.

From the general principle of work we can also arrive at this relation of effort and resistance to the grade. Suppose the weight W is rolled from the bottom to the top of the incline. Then it has been lifted H feet, and the work done is W (pounds) times H (feet), or WH foot pounds. But while the weight W has been traveling up the incline whose length is L, the force or pull P has moved down L feet, and the work put in is equal to P (pounds) times L (feet), or PL foot pounds. Therefore, if we neglect friction, we have

$$P \times L = W \times H,$$

or

$$\frac{P}{W} = \frac{H}{L}.$$

35. The grade of an incline. This ratio of the height to the length of an incline is expressed by engineers as so many feet rise per hundred feet along the incline, and is called the grade of the incline. For example, suppose a road rises

5 feet for every 100 feet along the incline, then this road is said to have a 5% **grade**. Since a 3% grade is the steepest allowable on a really good road, it is readily seen that a small force, such as can be exerted by a horse, can move a much heavier load up a gradual incline than could be lifted directly. For this reason the highways in mountain regions are laid out as zigzags and switchbacks. If we want a flight of steps easy to climb, we make the slope gentle.

Nevertheless it should be remembered that while the pull is less than the weight of the load, yet the distance the load travels is greater than when it is lifted straight up. In other words, what we gain in the amount of effort required we lose in the distance over which it must be exerted. The total work to be done is independent of the grade, except for the indirect effect of friction.

36. Wedge. If instead of pulling the load up the incline, we push the incline under the load, the inclined plane is called a **wedge**. Of course the smaller the angle of the wedge, the easier it is to drive it against the resistance. The fact that friction plays a very important part in its action makes it impossible to make a simple statement of the relation of the effort required to force in a wedge to the resistance to be overcome.

All cutting and piercing instruments, such as the ax, the chisel, and the carpenter's plane, as well as nails, pins, and needles, act like wedges. The carpenter uses wedges to fasten the heads of hammers and axes on their handles. The woodsman uses wedges to split logs of wood.

37. Screw. When an enormous force must be exerted, as in lifting a building, such machines as the lever, pulley, and inclined plane will not do, because we cannot get enough mechanical advantage. A screw, such as the **jack-screw** (Fig. 32), is sometimes used for this purpose. In one complete turn of the screw, the weight is lifted the distance between two successive threads, which is called the **pitch** of

the screw, while the effort is exerted through a distance equal to the circumference of the circle traced by the end of the bar or handle. In each complete turn the *output* is equal to the weight times the distance between two successive threads, and the *input* is equal to the effort times the distance through which it acts; namely, the circumference of a circle.


FIG. 32.—Jackscrew.

If W equals the weight to be lifted and p (pitch) equals the distance between threads, the output for one turn is W times p. Let P equal the effort or force applied on the handle, and $2\pi r$ equal the circumference of the circle in which it acts. Then P times $2\pi r$ is the input. Therefore applying the principle of work to the machine, we would have, if friction could be neglected,

$$W \times p = P \times 2\pi r,$$
$$\frac{W}{P} = \frac{2\pi r}{p}.$$

In other words, the *mechanical advantage of the screw is equal to the ratio of the circumference of the circle moved over by the end of the lever, to the distance between the threads of the screw.*

As a matter of fact, friction consumes a large part of the work put in, and therefore the input is greater than the output. But this loss is not wholly a disadvantage, for it keeps the screw from turning backward of itself.

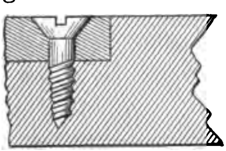


FIG. 33.—Wood screw.

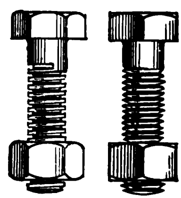


FIG. 34.—Bolts.

38. Applications of the screw. We are all familiar with carpenter's **wood screws** (Fig. 33) and machinist's **bolts** (Fig. 34). Ordinarily, however, we do not think of the

propeller of a boat or flying machine as a screw, but it is. The propeller (Fig. 35), with its two, three, or four blades fastened to one end of the shaft, is driven by an engine at the other end. Its rotation is so rapid that the water has no time to get out of the way, and the propeller screws itself through the water like a wood screw through wood.

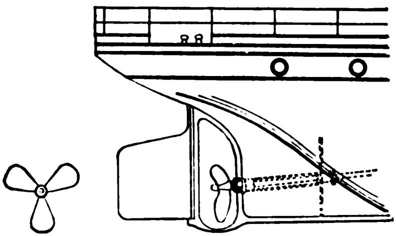


FIG. 35. — Screw propeller.

Another example of the screw is the **micrometer screw** (Fig. 36), which is used to make very precise measurements. It consists of an accurately turned thread of small pitch, perhaps 1 millimeter. It is evident that if such a screw is turned $\frac{1}{100}$ of a complete turn, the spindle moves along its axis just 0.01 millimeters. This is the easiest way of measuring so small a distance. In order to discover readily through just what fraction of a turn the screw is turned, the head is divided into 100 divisions.

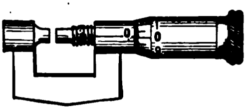


FIG. 36. — Micrometer screw.

39. Combinations of simple machines. What is called a single machine in factories and shops is usually a combination of the simple machine elements described above. It is, in fact, a more or less complicated collection of levers, pulleys, wheels, axles, and screws.

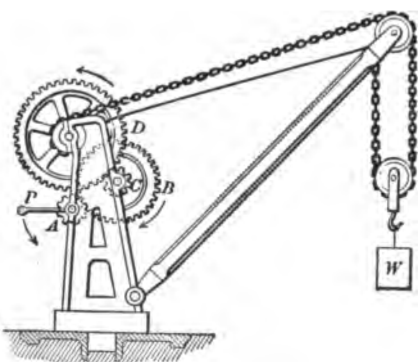


FIG. 37. — Builder's crane.

In order to show how such a machine may be analyzed

into its elements, let us, as it were, dissect a crane or derrick (Fig. 37) such as is used in unloading freight cars, or in hoisting building material into place.

The movable pulley to which W is attached gives a mechanical advantage of two; the fixed pulley at the end of the boom merely changes the direction of the pull; the wheel D and its axle give a mechanical advantage equal to the ratio of the size of the wheel to the size of the axle. A third mechanical advantage is gained in the wheel and axle, B and C, and finally there is the mechanical advantage of the crank P and the axle A. The total mechanical advantage of this compound machine is the product (*not* the sum) of the separate advantages gained by its separate elements. This is true of compound machines in general.

Problems

NOTE. Friction is to be neglected in these problems.

1. What force will be needed to pull a weight of 200 pounds slowly up a slope which rises 1 foot in 25 feet?

2. What weight can be moved on a 10 % grade with a pull of 50 pounds?

3. A boy, who can push with a force of 80 pounds, wants to roll a 200-pound barrel of flour into a cart 4 feet above the ground. How long a plank will he need?

4. What force is needed to move a 1500-pound wagon up a 3 % grade?

5. A test shows that it takes 1000 pounds more force to haul an electric car weighing 4 tons up a certain grade than to haul it along on a level. What is the grade?

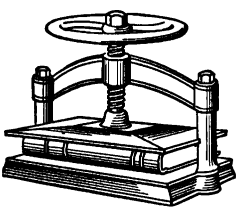


FIG. 38. — Letterpress.

6. What weight will be raised by a jack-screw when a force of 40 pounds is applied at the end of a lever arm 2 feet long, the pitch of the screw being 0.3 inches?

7. In a letterpress (Fig. 38) the threads are 0.25 inches apart and the hand wheel is 14 inches in diameter. If a 15-pound pull is applied to the rim of the wheel, how much force is brought to bear on the book?

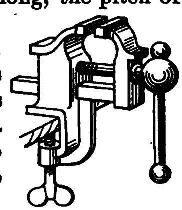


FIG. 39. — Machinist's vise.

8. The pitch of the screw of a bench vise (Fig. 39) is 0.2 inches and the handle of the screw is 7 inches long. What force could be exerted by the jaws of the vise if a force of 25 pounds were applied at the end of the handle?

9. The lever in a jackscrew extends 2 feet from the center. If a man is able to lift 25 tons by exerting a pressure of 100 pounds, how many threads to the inch must there be?

10. In the preceding problem, what is the mechanical advantage?

11. In the crane shown in figure 37, the weight W is 5 tons, and the radii of the three small cogwheels are supposed to be equal and each $\frac{1}{5}$ the radius of the crank P and of the wheels B and D, which are also equal. What is the mechanical advantage of the whole machine, and what force, neglecting friction, must be applied at P?

12. The pedal of a bicycle is halfway down and is pressed down with a force of 100 pounds. The crank arm is 6 inches long and the sprocket wheel is 8 inches in diameter. Find the tension or pull on the chain.

13. In the preceding problem the sprocket wheel attached to the rear wheel is 2.5 inches in diameter and the wheel is 28 inches in diameter. How far does the bicycle go when the pedal makes one complete revolution? How much does the tire of the rear driving wheel push backward on the roadbed when the man presses 100 pounds on the pedal?

40. Work and power. The words "work" and "power" are often confused or interchanged in colloquial use. The term "work" in physics means the overcoming of resistance. For example, if a boy carries a pail of water weighing 50 pounds up a flight of stairs 12 feet high, he does 600 foot pounds of work. The amount of *work* done would be the same whether he did this in one minute or one hour, but the amount of *power* required to do this job in one minute would be 60 times the power required to do it in one hour. The term "**power**" adds the notion of **time**. *Power means the speed or rate of doing work.*

41. Horse power. The earliest use of steam engines was to pump water from mines. This work had previously been done by horses; so the power of the various engines was estimated as equal to that of so many horses. Finally, James Watt carried out some experiments to determine how many foot pounds of work a horse could do in one minute. He found that a strong dray horse working for a short time could do work at the rate of *33,000 foot pounds per minute or*

550 foot pounds per second. This rate is therefore called a **horse power.** To get the horse power of an engine, compute the number of foot pounds of work done per minute and then divide by 33,000, or per second and divide by 550.

$$\textbf{Horse power (H. P.)} = \frac{\textbf{foot pounds per minute}}{\textbf{33000}} = \frac{\textbf{foot pounds per second}}{\textbf{550}}.$$

Suppose an engine is used to pump 10,000 gallons of water per hour into a reservoir 50 feet above the supply. How much horse power is required?

One gallon of water weighs 8.34 pounds; so 10,000 gallons of water weigh 83,400 pounds. The work done in lifting this weight 50 feet is 83,400 × 50, or 4,170,000, foot pounds. Since this is done in one hour, the work per minute is $\frac{4170000}{60}$ or 69,500 foot pounds. The horse power required would be $\frac{69500}{33000}$ or 2.1 H. P.

42. Transmission of power. In any shop containing several machines one easily distinguishes two kinds—the *driving machines*, which may be steam, gas or hot-air engines, or water or electric motors, and the *driven machines*, such as lathes, drills, planers, and saws. There must always be some connecting link between a driving and a working machine; that is, some means of *transmission.* If these machines are not far apart, the common method is to use shafting, belts, chains, or cogwheels; but when the prime mover and the driven machine are widely separated, sometimes even miles apart, some form of electrical transmission is used. Electrical transmission will be explained later in Chapter XVIII.

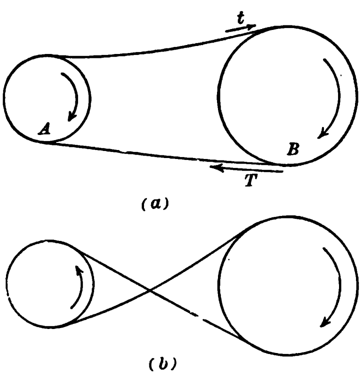


FIG. 40.—Transmission of power by a belt.

When a belt, rope, cable, or endless chain is used, it passes over two pulleys, as shown in figure 40. In case *a*, the pulleys rotate in the same direction,

while in case *b*, where the belt is crossed, they rotate in opposite directions. It is evident that the small pulley turns just as many times as fast as the large pulley, as the circumference (or diameter) of the small pulley is contained in the circumference (or diameter) of the large pulley.

The same is true of cogwheels, and since the teeth on the perimeters of two interlocking wheels must be the same size, it follows that the *number* of cogs on each wheel is a measure of its circumference. The speeds of two such wheels are inversely proportional to the number of teeth on them. Just as in the case of two pulleys with a crossed belt, two cogwheels rotate in opposite directions.

Suppose a pulley A is driving a second pulley B by means of a belt, as shown by the arrows in figure 40; both sides of the belt must be under some tension in order to give the necessary pressure on the pulleys, so that the friction may keep the belt from slipping. It is the usual practice to drive with the *upper side* of the belt slack, so that any sagging due to the weight of the belt may increase the arc of contact. The tension, then, on the lower side (T) must be greater than the tension on the upper side (t). It is the *difference in tension* ($T-t$) of the two sides of the belt which measures the *force* involved in the transmission of power. The work done in one minute is equal to the difference in tension times the speed of the belt in feet per minute.

$$\text{Horse power} = \frac{\text{difference in tension} \times \text{speed (ft. per min.)}}{33000}.$$

PROBLEMS

1. If it takes 22 pounds to pull a 200-pound sled along a level road covered with snow, how much work is done in dragging the sled 50 feet?

2. In the preceding problem, if the sled is drawn at the rate of 4 miles an hour, how many horse power are required?

3. How much work can a 5-horse-power engine do in 10 minutes?

4. What is the horse power of an elevator motor, if it can raise the car with its load, 1500 pounds in all, from the bottom to the top of a 100-foot building in 10 seconds?

5. An aëroplane with a 50 horse-power engine makes 60 miles an hour. What is the backward thrust of the propeller?

6. A locomotive pulling a train along a level track at the rate of 25 miles an hour expends 75 horse power. Find the total resistance overcome.

7. A motor has a 4-inch pulley which is belted to a 16-inch pulley on an overhead shaft. The motor is making 1800 revolutions per minute. What is the speed of the overhead shaft?

8. In an electric car motor a pinion or small cogwheel, attached to the armature shaft, has 20 cogs, and the gear wheel attached to the car axle has 36 cogs. If the car wheel is 33 inches in diameter, find the number of revolutions the motor makes while the car goes 100 feet.

9. If the tension T in the tight side of a belt 1 inch wide can safely be 44 pounds greater than the tension t in the slack side of the belt, how fast must the belt run to transmit 1 horse power?

10. It takes about 4 times as great a thrust to drive an aëroplane at 80 miles an hour, as to drive the same aëroplane at 40 miles an hour. Compare the horse powers required at the two speeds.

43. Friction. In the study of machines thus far we have assumed that we were dealing with *ideal* or *perfect* machines, in which the *output equals the input.* But in every actual machine the output is not quite equal to the input. This loss or waste of work is due to *friction.* By friction we mean the *resistance which opposes every effort to slide or roll one body over another.* This resistance, which always opposes the motion of the machine, depends on the condition of the rubbing surfaces. Great pains are therefore taken to diminish the friction as much as possible by making the surfaces which are to rub together smooth and hard, and by using various lubricants, such as soap and paraffin on wood, and grease, oil, and graphite on metal. For example, in a watch, the hardened steel axles turn in jewel bearings, which are the hardest and smoothest bearings known, and are lubricated with a special oil made for the purpose.

44. So-called laws of friction. The factors which control friction in any actual case are so numerous and so dependent upon the conditions, that only the most general principles may be stated positively. (1) Experience shows that starting friction is greater than sliding friction. For when we push a box across the table, we find that the force necessary to overcome the resistance of friction, which acts like a backward drag, is greater at the start than when the box is once in motion. (2) Friction does not much depend on velocity, but is a little greater at slow speeds. (3) Friction depends very much on the nature of the rubbing surfaces. (4) When a box is loaded, it requires much more force to pull it along than when it is empty. Careful experiments seem to show that the force needed to slide a given box over a certain floor is just about doubled when the pressure (weight of box and load) is doubled, and tripled when the pressure is tripled. That is, the force needed to overcome the friction seems to be proportional to the pressure. Experiments show that this force of friction may be a very small fraction of the pressure, such as 0.06 in the case of lubricated iron on bronze, or a large fraction of the pressure, such as 0.4 in the case of oak on oak without lubricant.

45. Coefficient of friction. *This fraction, the friction divided by the pressure, is called the coefficient of friction.*

$$\text{Coefficient of friction} = \frac{\text{force of friction}}{\text{pressure or weight}}.$$

In accordance with the statements in the last paragraph, the coefficient of friction for any particular pair of surfaces is pretty nearly constant for different loads or speeds. This is, however, only an approximation to the truth. Thus it has recently been found that the coefficient of friction of brake shoes on railroad car wheels nearly doubles when the speed drops from 60 miles an hour to 20 miles an hour. This is why an engineer or motorman lessens the pressure of his brakes as his train or car slows down.

The usefulness of even roughly accurate coefficients of friction is that they give some idea of how much resistance has to be overcome in any given case. Thus in the country, farmers often haul stones and pieces of heavy machinery on low sledges without wheels, called stone boats. To calculate how much force is needed to drag such a stone boat, one has only to look up the coefficient of friction between wood and dirt (about 0.66) in an engineer's handbook, and multiply it by the weight of the boat and its load. For

Force of friction = coefficient of friction × pressure.

46. Advantages of friction. In general it is true that friction reduces the amount of useful work which can be gotten out of a machine, yet it must not be forgotten that many machines depend upon friction for their operation. Without friction, belts would not cling to their pulleys, ropes could not be made, nails and screws would be useless, and even walking would be impossible, as any one can see who has experienced the difficulty of running on a polished floor or on ice. It is friction which has made possible our high-speed express trains; first because it is the friction or traction between the driving wheels of the locomotive and the rails that enables them to move at all, and second, because it is the friction between the brakes and the car wheels that enables them to stop quickly in case of emergency.

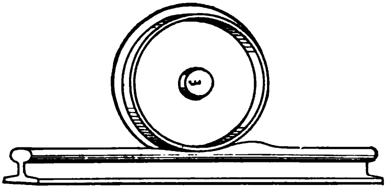


FIG. 41. — Rolling friction (exaggerated).

47. Rolling friction. Every one knows that the friction which opposes dragging a load along can be greatly reduced by mounting the load on wheels, but it should be noticed that even in this case there is something equivalent to friction between the wheels and the roadbed. When a car wheel rolls over a smooth track, as shown in

figure 41, its own weight and that of its load flatten it a little where it rests on the track, and also make a slight depression in the track. So as it rolls along it is continually forced to climb up out of the depression. Of course this depression is not easy to detect in the case of a steel track, but in the case of a soft dirt road it is very considerable. It is for this reason that the wheels of wagons which carry heavy loads are provided with wide tires so as to sink less into the roadbed; and for just this reason the hard surfaces of car wheels and tracks enable a locomotive to pull enormous loads. This resistance to rolling is called *rolling friction.*

The great advantage of ball and roller bearings is that they substitute rolling for sliding friction between the axles and their bearings. But even in ball bearings there is some sliding friction where adjoining balls rub against each other.

48. Efficiency of machines. The efficiency of a machine is the ratio of *output* to *input.* It is usually expressed as a per cent; that is, the output is a certain per cent of the input delivered to the machine.

$$\textbf{Efficiency} = \frac{\textbf{output}}{\textbf{input}} = \frac{\textbf{work done by machine}}{\textbf{work done on machine}},$$

or

$$\textbf{Output} = \textbf{efficiency} \times \textbf{input}.$$

For example, suppose we have an inclined plane of 5% grade (5 feet rise in 100 feet) and a load of one ton. If, because of friction, it takes a pull of 150 pounds to haul the load up the slope, what is the efficiency? In lifting 2000 pounds 5 feet, we do 10,000 foot pounds of work; this is the *output.* But we must pull with a force of 150 pounds through 100 feet, or *put in* 15,000 foot pounds of work. Therefore the efficiency is $\frac{10000}{15000}$, or 0.667, or 66.7%.

The efficiency of a lever where the friction is very small is nearly 100%, but in the commercial block and tackle it is sometimes less than 50%, and in the jackscrew, the friction is so large that the efficiency is often as low as 25%.

PROBLEMS

1. A tool is pressed on a grindstone with a force of 25 pounds; the coefficient of friction is 0.3. What is the backward pull of friction?

2. The coefficient of friction between the driving wheels of a locomotive and the rails is 0.25. How much must the locomotive weigh in order to exert a pull of 10 tons?

3. A test shows that it takes a pull of 17 pounds to pull on ice a man weighing 150 pounds. What is the coefficient of friction?

4. In lifting a 1250-pound block of marble to a height of 90 feet, the hoisting engine did 125,000 foot pounds of work. What was the efficiency of the hoist?

5. What load can a pair of horses, working at the rate of 2 horse power, draw along a level highway at the rate of 3 miles an hour, if the coefficient of friction of the wagon on the road is 0.17?

6. With a certain block and tackle it is found that a force of 125 pounds is necessary to lift a weight of 500 pounds, and the force must move 6 feet in order to raise the weight 1 foot. What is the efficiency of this block and tackle?

7. A motor whose efficiency is 90% delivers 5 horse power. What must be the input?

8. A hod carrier, weighing 160 pounds, carries 100 pounds of brick up a ladder to a height of 35 feet. How much work does he do in all? How much of it is useful work?

9. What is the efficiency of a pump which can deliver 250 cubic feet of water per minute to a height of 20 feet, if it takes a 10 horse-power engine to run it?

10. A steam shovel driven by a 6 horse-power engine lifts 200 tons of gravel to a height of 15 feet in an hour. How much work is done against friction?

SUMMARY OF PRINCIPLES IN CHAPTER II

The principle of moments: used in solving all kinds of levers, straight and bent, the wheel and axle, etc.; —

Effort × lever arm = resistance × lever arm.

To get force on fulcrum: or to solve a pulley system,

Sum of forces up = sum of forces down.

Laws of equilibrium: applicable to any object *at rest* under the action of two or more forces; —

(1) Sum of forces in any direction
= sum of forces in opposite direction.

(2) Sum of moments clockwise around any point
= sum of moments counter-clockwise around same point.

The principle of work; —

Work (foot pounds) = force (pounds) × distance (feet).

In any *frictionless* machine,

Input = output.

If there is *friction*,

Input = output + work lost by friction.

Power = rate of doing work.

1 horse power = 550 foot pounds per second,
= 33,000 foot pounds per minute.

$$\textbf{Coefficient of friction} = \frac{\textbf{force of friction}}{\textbf{pressure}}.$$

Force of friction = coefficient × pressure.

$$\textbf{Efficiency} = \frac{\textbf{output}}{\textbf{input}}.$$

Output = efficiency × input.

QUESTIONS

1. Make a list of a dozen applications of the simple machine elements described in this chapter that you have seen outside of the classroom within a week.

2. Distinguish between the popular use of the term "work" and its technical use in physics and engineering. Give an example of "work" that is not technically "work."

3. Analyze the working of the following machines: clothes wringer, broom, ice-cream freezer, plow, grindstone, and rotary meat chopper.

4. Distinguish between the terms "mechanical advantage" and "efficiency." Illustrate by an example.

5. Is there any "mechanical advantage" in an equal arm lever? Why is it often used in machines?

6. Why is an unequal arm lever useful?

7. Show how the principle of work applies to the lever.

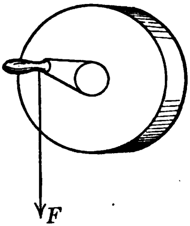


FIG. 42. — Crank on grindstone.

8. How would you calculate the moment of the force F as applied to the grindstone in figure 42?

9. Why are you likely to twist off the head of a screw by using a screwdriver in a bit brace?

10. When a machinist speaks of "an 8–32 screw," what does he mean?

11. What is meant by a perpetual-motion machine?

12. What kind of lubricant is used on journals of car wheels? What kind on clocks and watches? Why the difference in kind of lubricant?

13. What determines the "angle of repose," or slope, of the rock waste, or talus, at the base of a cliff?

14. Why are the modern air brakes on cars more effective than the old-fashioned hand brakes?

CHAPTER III

MECHANICS OF LIQUIDS

Hydraulic machines — Pascal's principle of transmitted pressure — applications in presses and elevators — pressure in a liquid due to its weight — levels of liquids in connecting vessels — upward pressure of liquids — Archimedes' principle and its applications — specific gravity of solids and liquids — city water works, faucets, gauges, and meters — water wheels — interaction between solids and liquids — capillarity.

49. Hydraulic machines. As we continue our study of machines we find some machines that involve more than the simple elements, the lever, the pulley, and the screw. For instance, there are a great many machines that make use of liquids, such as the water wheel, hydraulic press, and hydraulic elevator. These are called hydraulic machines, and this chapter will be devoted to the study of them. In the course of it we shall also have to consider dams and reservoirs, as well as all sorts of things that float or sink in water.

50. Pressure transmitted by a liquid. Suppose we fill a bottle with water and close it with a one-hole rubber stopper. Then let us fasten the stopper securely, as shown in figure 43, and force into the hole a metal rod of such a size as to fit rather tightly. The force applied to the rod will be transmitted to the inner surface of the bottle, and the bottle will burst.

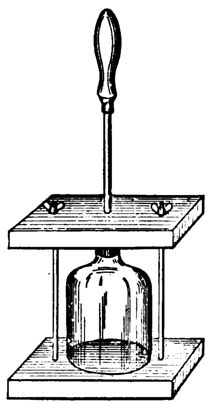


FIG. 43. — Water transmits pressure of piston.

This experiment shows that the water which is pressing against the bottom of

the piston is also pressing against everything else that it touches.

51. Pascal's principle. It seems reasonable to suppose that if we had a box filled with water and fitted with two equal pistons A and B, as in figure 44, the water would press equally hard on each piston. It is also evident that if a third equal piston were placed in the side of the box, as at C, the water would press sideways on it with an equal force.* In short, *if a liquid is pressing against any square inch with a certain force, it is pressing equally hard against every square inch of everything it touches.*

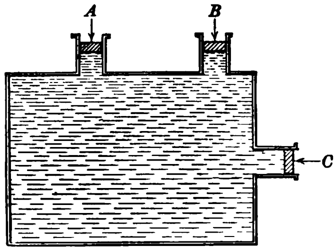


FIG. 44. — Pascal's principle.

52. The hydraulic press. The most useful application of this principle can be described in Pascal's (1623–1662) own words: "If a vessel full of water, closed in all parts, has two openings, of which the one is a hundred times the other, placing in each a piston which fits it, the man pushing the small piston will equal the force of a hundred men who push that which is a hundred times as large, and surpass that of ninety-nine. Whatever proportion these openings have, and whatever direction the pistons have, if the forces that apply on the pistons are as the openings, they will be in equilibrium."

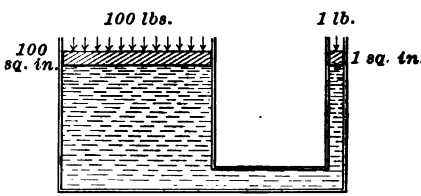


FIG. 45. — Diagram of hydraulic press.

This refers to a mechanism like that shown in figure 45. Suppose there is a force of one pound pushing down on the small piston, and that the large piston has 100 times as great an area. Then there must be 100 pounds pushing down on the large piston to balance it. It will be seen, however,

* The effect of the weight of the liquid is here neglected.

Blaise Pascal. French scientist and mathematician. Born 1623. Died 1662. Studied the pressures exerted by liquids and gases. Famous also for his achievements in mathematics.

that the pressure on each square inch of the large piston is one pound. In other words the pressure has been transmitted by the liquid so as to act with the same force on every square inch.

53. Applications of the hydraulic press. This device of Pascal gives us an easy way of exerting enormous forces, such as are needed in baling paper, cotton, etc., in punching holes through steel plates, and for extracting oil out of seeds. The commercial machine (Fig. 46) is exactly like that described by Pascal except that there is usually a check valve (*v*) between the small piston and the big one, and the small piston is arranged to work like a pump, with a valve (*d*) at the bottom for admitting more oil. Often the small piston is forced down by a lever. The method of operation is simple. On the upstroke of the pump piston, the valve at the bottom of the pump opens and oil flows in from the reservoir. On the downstroke of the pump piston, the oil is forced over through the connecting pipe past the valve, and pushes the large working piston up very slightly. If the large piston is 100 times as large in cross section as the small piston (*i.e.* diameters as 10 : 1), the large piston is lifted only $\frac{1}{100}$ the distance the pump piston is pushed down each stroke. But since the force exerted by the large piston is, neglecting friction, 100 times that applied to the small piston, it follows that the work done on the machine is equal to the work done by the machine. If we consider the work done against friction, the equation becomes, —

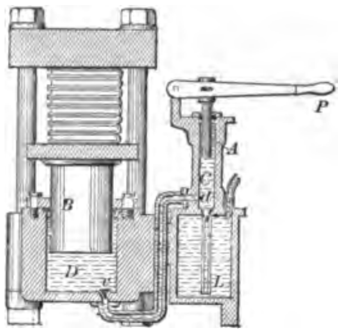


FIG. 46. — Hydraulic press.

Input = output + work done against friction.

54. Working model of a hydraulic press. Let us try to appreciate the tremendous forces which are obtainable with the hydraulic press by

operating a model press, such as is shown in figure 47, to break a stick of wood. By measuring the diameters of the pistons, and the lengths of the lever arms, we may calculate the total mechanical advantage of the machine.

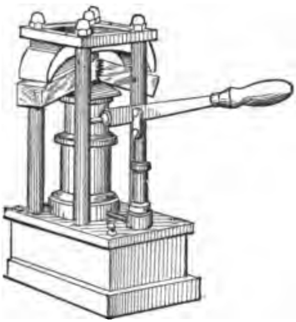


FIG. 47.—Working model of hydraulic press.

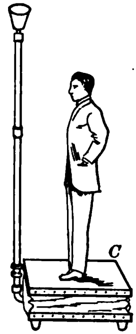


FIG. 48.—Hydrostatic bellows.

Another striking experiment is to let a boy balance his own weight against a column of water by means of the hydrostatic bellows (Fig. 48). By calculating the actual areas involved and the force acting on each square inch, we may compute the height of water that should be required and compare this with the actual height.

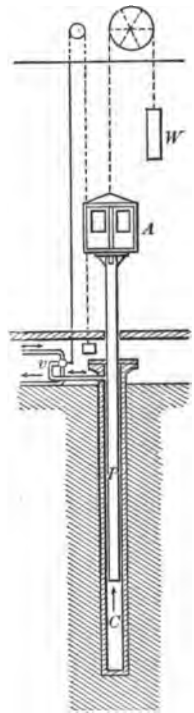


FIG. 49.—Hydraulic elevator.

55. Hydraulic elevators. Pascal's principle is also used in hydraulic elevators, which are commonly employed where heavy machinery is to be lifted. A simple form is shown in figure 49. At the bottom of the elevator well is a pit as deep as the building is high. In the pit is a cylinder (C) and in this cylinder is a plunger (P), to the top of which the elevator cage (A) is firmly fastened. When water under pressure (often simply the pressure of the water mains) is admitted through the valve (v) into the cylinder, the plunger rises and forces up the elevator. The weight of the elevator is partly counter-

balanced by a weight (W). When the operator, by pulling the cord, turns the valve so as to connect the cylinder in the pit with the sewer pipe, the elevator comes down.

When speed is demanded, as in high office buildings, the motion of the hydraulic plunger is communicated to the cage by a cable passing over a series of pulleys, so that the cage moves four times as far and four times as fast as the plunger.

56. Pressure and force. It is necessary to distinguish between the terms **pressure** and **force.** *Force means a push or a pull,* and is usually expressed in terms of the push or pull necessary to hold up a given weight, such as a pound or a kilogram. *Pressure means the push or pull per unit area of surface.* Pressure may be expressed in various ways, for example, as so many grams per square centimeter or so many pounds per square inch.

The real advantage of the hydraulic press is that, although the *pressure* on the large piston is exactly the same as that on the small piston, the *force* exerted by the large piston is many times greater.

PROBLEMS

1. If the diameters of two pistons in a hydraulic press are 1 inch and 10 inches, what are their areas of cross section?

2. If the small piston in problem 1 is subjected to a pressure of 10 pounds per square inch, what pressure, neglecting friction, must be applied to the large piston to hold it in place?

3. If a total force of 10 pounds is applied to the small piston in problem 1, what total force must be applied to the large piston to hold it in place?

4. The diameters of the pistons in a hydraulic press are 20 inches and 1 inch. What must be the force on the small piston if a force of 5 tons is to be exerted by the large piston?

5. In problem 4, suppose the small piston to move 1 foot. How far does the large piston move?

6. If the water pressure in a city water main is 50 pounds per square inch and the diameter of the plunger of an elevator is 10 inches, how heavy a load can the elevator lift? If the friction loss is 25%, what load can be lifted?

57. Pressure in a liquid due to its weight. Not only does a liquid transmit pressure when it is in a closed vessel, but a liquid in an open vessel, such as water in a tin pail, exerts a pressure on the bottom of the vessel because the liquid is itself heavy. This **bottom pressure**, that is, the force on each square inch, evidently *depends on the depth of the liquid*, and also *on its density*.

For example, suppose we have a box with a bottom 10 centimeters by 20 centimeters and 15 centimeters deep, filled with water. Then on each square centimeter of the bottom of the box there rests a column of water 15 centimeters tall, weighing 15 grams, and so the pressure on the bottom is 15 grams per square centimeter. The total downward force of the water against the bottom would be 200×15, or 3000 grams, for

$$\text{Total force} = \text{area} \times \text{pressure}.$$

If the box were filled with mercury instead of water, the pressure on the bottom would be the weight of a column of mercury 15 centimeters high and 1 square centimeter at the base; that is, the weight of 15 cubic centimeters of mercury. Since 1 cubic centimeter of mercury weighs 13.6 grams, 15 cubic centimeters would weigh 15×13.6, or 204 grams. The total force of the mercury pushing down on the bottom of the box would be 200×204, or 40,800 grams, or 40.8 kilograms.

58. Bottom pressure and shape of vessel. So far we have considered vessels with vertical sides such as *A* in figure 50.

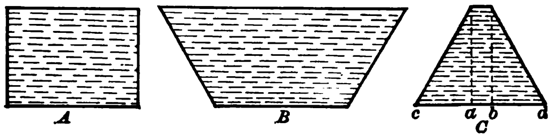


FIG. 50. — Vessels with (*A*) vertical, (*B*) flaring, and (*C*) conical sides.

In the ordinary pail, however, the sides are not vertical, but flare outward as shown in *B* in figure 50. Perhaps one might expect that the pressure on each square centimeter of the bottom would be greater than in case *A*, because there is so much more water in the vessel. This, however, is not the case. Each square centimeter of the bottom has to hold

up only the little column of water above it just as it did in case *A*. The extra water above the slanting sides is held up by those sides and not by the bottom. If the area of the base and depth of liquid is the same in both *A* and *B*, then the total downward push of the liquid on the bottom will be the same even though *B* holds more liquid than *A*.

In case *C*, the depth of liquid and area of base are the same as in cases *A* and *B*, but the top is smaller than the base. It is easy to see that the pressure on that portion of the base *ab* directly under the top would be the same as in the other vessels, but it might at first seem that the pressure would gradually decrease as we go from *a* to *c* and from *b* to *d*. There is an interesting experiment devised to settle this question.

59. Experiments with Pascal's vases. The apparatus (Fig. 51) consists of three glass vessels of shapes to correspond roughly to *A*, *B*, and *C* in figure 50. The bottom of each vessel is made the same size and screws into a short cylinder, across the bottom of which is tied a disk of sheet rubber. The pointer below is a lever with its short arm pressing against the center of the rubber disk, and the long arm moves up and down across a scale.

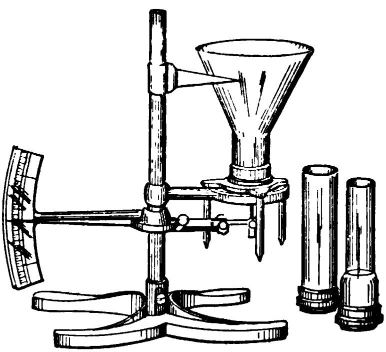


FIG. 51. — Pascal's vases.

With this apparatus it is possible to show that (*a*) *the downward pressure of a liquid is proportional to the depth*, (*b*) *the downward pressure of a liquid is proportional to its density, and* (*c*) *the downward pressure in a liquid is independent of the shape of the vessel.*

It seems impossible that unequal quantities of water should exert an equal downward push against the bottom. But if we recall that when the sides slope outward, the sides

hold up the excess of water, we can see that when the sides slope inward, they push down enough to make up for the deficit in water.

60. Liquids also exert pressure sidewise. We all know that if a hole is bored in the side of a tank or barrel of water, the water will spurt out. This means that before the hole was bored the liquid must have been pressing against that bit of the side of the barrel. Liquids, then, exert a sidewise pressure due to their weight, as well as a downward pressure.

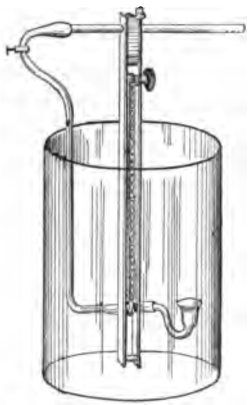


FIG. 52. — Pressure gauge to show pressure equal in all directions.

We can investigate how this sidewise pressure varies with the depth and the direction by means of the gauge shown in figure 52. The apparatus consists of a rubber diaphragm, which may be turned about a horizontal axis, and is connected by a rubber tube to a horizontal glass tube containing a globule of some colored liquid. As we lower the pressure gauge into the jar of water, we observe that the globule moves to the right showing a gradual increase of pressure with increase of depth. If we repeat this with the diaphragm facing in another direction, we get the same result. If we hold the frame at some fixed depth, and rotate the diaphragm around a horizontal axis, we find the globule remains practically stationary, showing that the pressure is the same in all directions.

*The **sidewise pressure** of a liquid increases with the depth and density of the liquid. At a given depth a liquid presses downward and sidewise with exactly the same force.*

61. Calculation of sidewise pressure. To calculate the sidewise push of water against a dike or dam, we have to remember that both the downward and sidewise pressure increase gradually from zero at the surface to their value at the bottom. We have already seen that this bottom pressure is equal to the weight of a column of water with a base one unit square and with a height equal to the depth. The

average sidewise pressure is equal to the pressure halfway down, or is one half the bottom pressure. The total sidewise push of the water against the dam is then equal to the area times the average pressure.

For example, suppose we have a box 10 centimeters wide, 20 centimeters long, and 15 centimeters deep filled with water. What is the *total force* tending to push out the *end* of the box? The pressure at a point halfway down the side would be 7.5 grams per square centimeter. There are in the end 10 × 15, or 150 square centimeters. Therefore the total force against the end is 150 × 7.5, or 1125 grams.

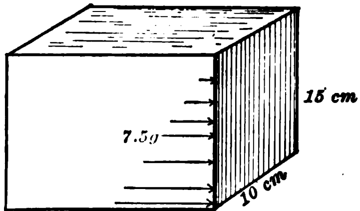


FIG. 53. — Sidewise push of water against end of box.

Again, suppose the box were a large tank full of water, and the dimensions, expressed in feet, were 10 by 20 by 15. What is the end thrust? The pressure halfway down would be the weight of a column of water with 1 square foot for its base and 7.5 feet high, *i.e.* 7.5 × 62.4, or 468 pounds per square foot. Since there are 10 × 15, or 150 square feet, in the end of the tank, the total end thrust is 150 × 468, or 70,200 pounds, or about 35 tons.

62. Levels of liquids in connecting vessels. Probably every one has observed that water stands at the same level in the spout of a teakettle as in the kettle itself (Fig. 54). In other words, **liquids seek their own level,** or the same liquid in any number of connecting vessels will have its free surface at the same level in each. This is to be expected from the fact that the pressure in a liquid depends upon the depth below the free surface. Thus if any point in the connecting portion between the two vessels were unequally far below the two surfaces, the pressures in either direction would not

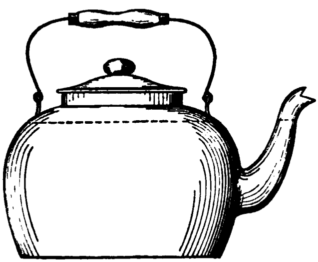


FIG. 54. — Water seeks its own level in a teakettle.

balance, and the liquid would flow from one vessel to the other until the levels were equalized.

The water gauge on a steam boiler (Fig. 55) is a good application of this principle. The gauge consists of a thick-walled glass tube which connects at the top with the steam space, and at the bottom with the water in the boiler. The valves A and B are closed when the glass tube is to be replaced. The valve C is opened occasionally to test the gauge to see that it reads correctly and has not clogged up.

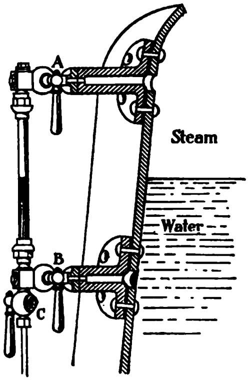


FIG. 55. — Water gauge on a boiler.

63. Upward pressure of liquids. If one tries to push a pail under water bottom downward, he finds he must overcome considerable resistance because of the upward push of the water on the pail. In order to see just how much this upward push of the water is, let us try the following experiment.

Let a glass cylinder, which has its bottom edge ground off smooth, be closed with a glass plate or piece of cardboard, held in place by a thread, as shown in figure 56. When we push this cylinder into a jar of water, we can let go the thread and yet the glass bottom will not fall off. It is evident that there is an upward pressure due to the water, and the next question is, how much? If we pour colored water into the cylinder until the bottom drops off, we shall have to fill the cylinder until the levels inside and outside are the same.


FIG. 56. — Upward pressure of water.

In general we may say that *the upward pressure exerted by a liquid at any depth is equal to the down-*

ward pressure which would be exerted by the same liquid at the same depth.

PROBLEMS

1. The water in a standpipe is 10 meters deep. What is the pressure on one square centimeter of the bottom?

2. The water in a standpipe is 40 feet deep. What is the pressure on one square inch of the bottom?

3. If the diameter of the tank in problem 2 is 10 feet, what is the total force which the bottom of the tank must sustain?

4. A diver goes down into sea water (density 1.03 grams per cubic centimeter) to a depth of 10 meters. What is the pressure on him in kilograms per square centimeter?

5. The hydraulic engineer speaks of pressure as "head of water," which means the pressure due to the weight of column of water as high as the "head of water." Express in pounds per square inch a "head of 50 feet."

6. What is the pressure, near the keel, on a vessel drawing 6 meters?

7. Figure 57 is a cylindrical tank 10 × 12 centimeters; out of the top rises a tube 20 centimeters long. The box and tube are filled with water.

(*a*) Find the pressure in grams per square centimeter at the bottom of the tank.

(*b*) Does the size of the tube affect the pressure on the bottom?

(*c*) Find the pressure halfway up the side of the tank.

(*d*) Find the pressure at the top of the tank.

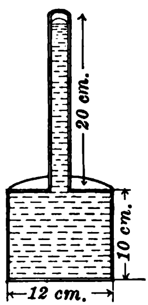


FIG. 57. — Box and tube full of water.

8. A rectangular tank is 5 feet wide, 10 feet long, and 4 feet deep. Calculate the total force exerted on the end when the tank is full of water.

9. Assuming that a cubic inch of mercury weighs 0.49 pounds, find the pressure on the bottom of a tumbler in which the mercury stands 4 inches deep.

10. How high a column of water could be supported by a pressure of one kilogram per square centimeter?

11. If the density of mercury is 13.6 grams per cubic centimeter, what is the pressure exerted at the base of a column 76 centimeters high?

12. A dam is 50 feet long and 6 feet high, and the water just reaches the top. What is the total force against the dam?

13. A hole 6 inches square is cut in the bottom of a ship drawing 18 feet of water. What force must be exerted to hold a board tightly against the inside of the hole?

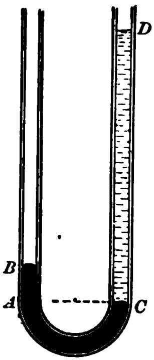


FIG. 58. — U-tube with mercury and another liquid.

14. How much "head of water" is needed to give a pressure of 1 pound per square inch?

15. What must be the difference in height between a fire hydrant and the surface of the water in a city reservoir to give a pressure of 50 pounds per square inch at the hydrant?

16. In figure 58, the U-tube is partly filled (BAC) with mercury whose density is 13.6 grams per cubic centimeter, and partly (CD) with a liquid of unknown density. If the length of the column BA is 5 centimeters and that of the column CD is 75 centimeters, what is the density of the liquid?

64. Buoyant effect of liquids. When swimming in deep water, we find that our bodies are very nearly floated. When we pick up a stone under water, we find it much heavier if we lift it above the surface. Things seem to be lighter under water; in other words, water buoys up anything placed in it. In order to find how much lighter anything is under water than it is out of water let us try the following experiment.

We have a hollow metal cylindrical cup C, and a cylindrical block B, which has been nicely turned to fit inside the cup C. We hang both from a beam balance, as shown in figure 59, and counterbalance with a weight W on the other scalepan. Then we bring a glass of water up under the block B, so that it is entirely under water. The left-hand side of the balances rises, which shows the upward push of the water upon B. But we can restore the equilibrium again by pouring water into the cup C until it is just filled. This shows that B *loses in apparent weight the weight of its own bulk of water.* If we try the experiment, using kerosene instead of water, we find that exactly the same thing is true.

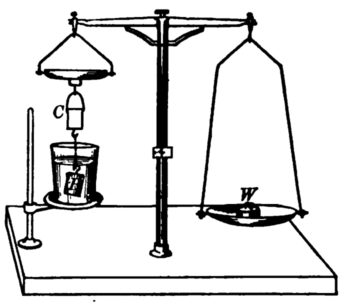


FIG. 59. — Buoyant effect of liquids.

65. Archimedes' principle. The principle proved by this experiment may be stated as follows: —

The loss of weight of a body submerged in a liquid is the weight of the displaced liquid.

It is supposed that this principle about the loss of weight of a body in a liquid was discovered by the old Greek philosopher Archimedes (287–212 B.C.). Hiero, king of Syracuse, suspected a goldsmith who had made a crown for him, and ordered Archimedes to find out if any silver had been mixed with the gold in the crown. To do this without destroying the crown seemed a puzzle at first, but one day, while Archimedes was in the public bath, he noticed that his body was buoyed up by the water in which it was submerged. Seeing in this effect the solution of his problem, he leaped from the bath and rushed home shouting, "Eureka! Eureka!" (I have found it! I have found it!).

66. Explanation of Archimedes' principle. This principle will be readily understood from the following example. Suppose we place a rectangular block in a jar of water, as shown in figure 60. Let the block be $10 \times 6 \times 4$ centimeters and let the top be 5 centimeters below the surface of the water, and the bottom 15 centimeters beneath the surface. Then the pressure on top, that is, the downward push on each square centimeter, is 5 grams and the pressure on the bottom, that is, the upward push on each square centimeter, is 15 grams. Since the top and bottom each have an area of 6×4, or 24 square centimeters, the whole upward push on the bottom is 24×15, or 360 grams, while the whole downward push on the top is only 24×5, or 120 grams. This leaves a net upward force or buoyancy of 240 grams. But this is exactly the weight of the displaced water, for the volume of the displaced water is $10 \times 6 \times 4 = 240$ cubic centimeters, and we have seen in section 11 that this much water weighs 240 grams.

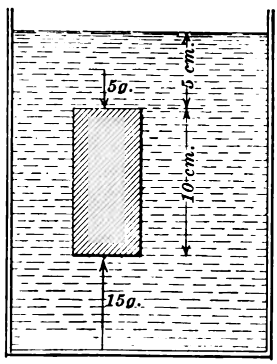


FIG. 60. — Lifting effect of water on submerged block.

The same sort of reasoning would hold at any depth and for any liquid other than water and with any irregular-shaped body. So it may be said that in any liquid of any density a body seems lighter by the weight of the displaced volume of *that liquid.*

67. Floating bodies. Let us think what will happen if this upward force, or buoyant force, is more than the weight of the body submerged. Evidently the body will rise and will continue to rise as long as the upward push remains greater than the downward pull of gravity. But as soon as any of the body projects above the surface, less water is displaced and the upward push is less. When enough of the body projects to reduce the buoyant force to equality with the weight, the body stops rising and floats. In this case we see that the loss of weight is the whole weight itself.

A floating body displaces its own weight of the liquid it is floating in.

The following experiment will help to make this principle of Archimedes, as applied to floating bodies, seem more real. Suppose we balance an overflow can on a platform scale, as shown in figure 61. The can is filled with water so that it just overflows and is balanced by the weight on the other platform. We will place a dish to catch the overflowing water, and then put a block of wood gently in the can. After the water has stopped overflowing, it will be seen that the scales again balance. This means the weight of water which flowed over was just equal to the weight of the block. This can be verified in another way by weighing the water displaced by the block.

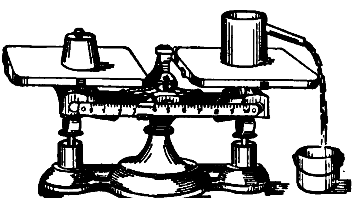


FIG. 61.—Weight of liquid displaced by floating body.

68. Applications of Archimedes' principle. If we know the total weight of a ship and its equipment, we can tell at once what weight of water it will displace, and so it is possible to compute how deep it must sink to displace its own weight

of water. It is also evident that a boat must sink a little deeper in fresh water than in salt water, and will sink deeper when loaded than when empty. A submarine boat is so constructed that it is only slightly lighter than water. It can then be submerged by letting water into certain tanks and can be made to rise by pumping the water out of the tanks. This same idea is made use of in the floating dry-dock shown in figure 62. When the tanks *T*, *T*, *T* are full of water, the dock sinks until the water level is at *LL*. The ship to be repaired is then floated into the dock and the water is pumped out of the tanks *T*, *T*, *T*. As the compartments are emptied of water, the dock rises until the water level is at the line *W*, *W*, lifting the ship out of water. The ship and dry-dock still displace their own weight of water, but the displacement is in a different place.

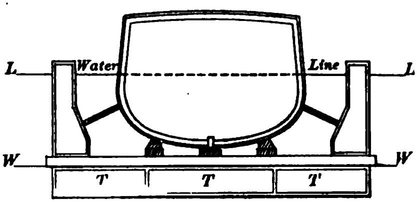


FIG. 62.—Floating dry-dock.

PROBLEMS

1. A piece of stone weighing 235 grams in air and 128 grams in water is put into a dish just full of water. How much water runs over?

2. A rowboat weighs 200 pounds. How many cubic feet of water does it displace?

3. A barge is 30 feet long and 16 feet wide, and has vertical sides. When a large elephant is driven on board, it sinks 4 inches farther in the water. How many tons does the elephant weigh?

4. What is the volume of a 125-pound boy, if he can float entirely submerged except his nose?

5. A rectangular block is 22 centimeters long, 6 centimeters wide, and 4 centimeters high, and floats in water with 1 centimeter of its height above water. How much does it weigh?

6. A cube 5 centimeters on an edge weighs 600 grams in air. How much does it weigh in water?

7. How much will a cubic foot of brass (density 8.4 grams per cubic centimeter) weigh in gasolene (density 0.79 grams per cubic centimeter)?

8. A rectangular solid $10 \times 8 \times 6$ centimeters is submerged in water, so that the top, whose dimensions are 10×8 centimeters, is horizontal and 12 centimeters below the water surface.

(a) Find the total force pressing down on the top.
(b) Find the total force pushing up on the bottom.
(c) Find the loss of weight of the solid.

69. Specific gravity and density. Archimedes' principle furnishes us with a convenient method of comparing the weight of a substance with the weight of an equal bulk of water. The ratio of these weights is called the *specific gravity* of the body. In other words,

$$\text{Specific gravity} = \frac{\text{weight of body}}{\text{weight of equal bulk of water}}.$$

For example, a piece of marble weighs 100 grams and an equal bulk of water weighs 40 grams, then the marble is 100/40 or 2.5 times as heavy as the water. The specific gravity of marble, then, is 2.5.

The term **specific gravity** does not mean quite the same thing as **density**. The specific gravity of a substance is an *abstract* number; for example, the specific gravity of mercury is 13.6. But the density of a substance is a *concrete* number; for example, the density of mercury is 13.6 grams per cubic centimeter, or 850 pounds per cubic foot.

In the metric system, the density of water is one gram per cubic centimeter, and therefore

$$\textbf{Density (g. per cm.}^3\textbf{)} = \textbf{(numerically) specific gravity.}$$

In the English system, the density of water is 62.4 pounds per cubic foot, and therefore

$$\textbf{Density (lbs. per cu. ft.)} = \textbf{(numerically) } 62.4 \times \textbf{specific gravity.}$$

70. Methods of determining specific gravity of solids.

GENERAL RULE. *First weigh the object. Next find by some indirect method the weight of an equal bulk of water. Finally divide the weight of the object by the weight of the equal bulk of water.*

This general statement covers all the various processes for finding the specific gravity either of solids or of liquids. The different procedures vary only in the method of finding the weight of an equal bulk of water.

1st Method. If the object is a **regular geometrical solid,** you can measure its dimensions and **calculate its volume,** and from that get the weight of an equal bulk of water.

2d Method. If the object is a solid that will **sink in water,** and will not dissolve, you can determine its **loss of apparent weight in water.** This is the weight of an equal bulk of water. That is,

$$\textbf{Specific gravity} = \frac{\textbf{weight of body}}{\textbf{loss of weight in water}}.$$

For example, suppose a piece of copper weighs 178 grams in air and 158 grams in water. The loss, 20 grams, is the weight of an equal bulk of water. Therefore the specific gravity of copper = 178/20 = 8.9.

3d Method. If the object is **lighter than water,** and does not dissolve, select a sufficiently large **sinker** and suspend it below the object, as shown in figure 63. Then bring a jar of water up under the whole thing until the water level is between the sinker and the object, and weigh. Then raise the jar still farther until the water level is above the object, and weigh again. This weight will be less than the first because in this case the water buoys up the object, while in the first case it does not. The difference between the two weights is equal to the weight of the water displaced by the object.

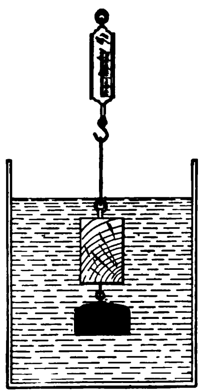


FIG. 63.—Specific gravity with sinker.

$$\textbf{Specific gravity} = \frac{\textbf{weight of body}}{\textbf{lifting effect of water on body only}}.$$

It will be noticed that in this case the loss of weight or lifting effect of the water on the body is larger than the whole weight. This is why the body floats.

For example, suppose a piece of wood weighs 120 grams in air, and that, with a suitable sinker, it weighs 270 grams when the sinker is under water, and 90 grams when both are under water. Then the lifting effect of the water on the wood is 270 – 90, or 180 grams. Therefore the specific gravity of the wood is 120/180 = 0.667.

71. Specific gravity of liquids.

1st Method. Weigh a **bottle** empty, then full of the liquid, and then full of water. Subtract the weight of the empty bottle in each case, and then compare the weight of the liquid with the weight of an equal volume of water.

$$\text{Specific gravity} = \frac{\text{weight of liquid}}{\text{weight of equal volume of water}}.$$

Bottles, called **specific gravity flasks** (Fig. 64), are made for the purpose of determining the specific gravity of liquids with great accuracy and facility. They are usually made to contain a definite quantity of pure water at a specified temperature; for example, 250 grams.

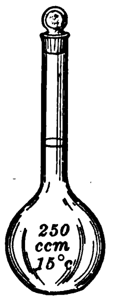


FIG. 64. — Specific gravity flask.

2d Method. Weigh a piece of glass in air, then in the liquid, and then in water. Find the **loss of weight in the liquid** and the **loss of weight in water.** This loss of weight in the liquid is the weight of the liquid displaced, and the loss of weight in water is the weight of an equal volume of water. Then

$$\text{Specific gravity} = \frac{\text{loss of weight in liquid}}{\text{loss of weight in water}}.$$

For example, suppose the glass weighs 330 grams in air, 150 grams in sulphuric acid, and 230 grams in water. The glass loses 180 grams in acid and 100 grams in water. Since these are the weights of equal volumes of acid and water, the specific gravity of the acid = 180/100 = 1.8.

*3d Method.** The most common way of determining the specific gravity of liquids is by the **hydrometer.** This is usually made of glass, and consists of a cylindrical stem and a bulb weighted with mercury or shot to make it float upright (Fig. 65). The liquid is poured into a tall jar, and the hydrometer is gently lowered into the liquid until it floats freely. The point where the surface of the liquid touches the stem of the hydrometer is noted. There is usually a paper scale inclosed inside the stem, so made that the specific gravity (or density in grams per cubic centimeter) can be read off directly. In light liquids, like kerosene, gasolene, and alcohol, the hydrometer must sink deeper to displace its weight of liquid than in heavy liquids like brine, milk, and acids. In fact it is usual to have two separate instruments, one for heavy liquids, on which the mark 1.000 for water is near the top, and one for light liquids, on which the mark 1.000 is near the bottom of the stem.

FIG. 65. Hydrometer.

72. Commercial uses of the hydrometer. Since the commercial value of many liquids, such as sugar solutions, sulphuric acid, alcohol, and the like, depends directly on the specific gravity, there is extensive use for hydrometers. Perhaps the best-known form of hydrometer is the kind used in testing milk, called a **lactometer.** The specific gravity of cow's milk varies from 1.027 to 1.035. Since only the last two figures are important, the scale of a lactometer is made to run from 20 to 40, which means from 1.020 to 1.040. The specific gravity of milk does not give us a conclusive test as to its worth. Milk contains besides the water

* There is another method, using balancing columns, which will be described in the Laboratory Manual. To understand it one must have read Chapter IV.

(which is about 87 %) some substances which are heavier than water, such as albumen, sugar, and salt, and others that are lighter than water, such as butter fat. Besides the specific gravity, one needs to determine the amount of fat, and, if possible, the other solids in the milk, in order to know its richness. Of course the very important question as to the cleanliness of milk must be left to the bacteriologist.

PROBLEMS

1. A piece of ore weighs 42 grams in air and 25 grams in water. Calculate its specific gravity.

2. A stone weighs 15 pounds in air and 9 pounds in water.
(*a*) Find its specific gravity.
(*b*) Find its density in the metric system.
(*c*) Find its density in the English system.

3. A body has a specific gravity of 3.5. What is its density in (*a*) the metric system, and (*b*) the English system?

4. If the specific gravity of lead is 11.4, how many cubic centimeters of lead does it take to make a kilogram weight?

5. If the specific gravity of cork is 0.25, how many cubic feet of cork are there in 1 pound of cork?

6. A block of wood, 15 × 10 × 8 centimeters, floats with one of its largest sides 2 centimeters out of water.
(*a*) Find its weight.
(*b*) Find its specific gravity.

7. A plank 8 centimeters thick floats with 5 centimeters under water. Find its specific gravity.

8. A block of wood weighs 150 grams; a sinker is suspended from it, and when the sinker is under water and the block is in air, the combination weighs 350 grams. When the wood and the sinker are both under water, they weigh 100 grams. Find (*a*) the volume of the block of wood, and (*b*) its specific gravity.

9. A cube of iron 10 centimeters on an edge (specific gravity 7.5) floats in mercury (specific gravity 13.6). How many cubic centimeters are above the mercury?

10. A can weighs 190 grams when empty, 600 grams when full of water, and 613 grams when full of milk.
(*a*) What is the capacity of the can in cubic centimeters?
(*b*) What is the specific gravity of the milk?

11. How much does 1 cubic centimeter of lead (specific gravity 11.4) weigh in kerosene (specific gravity 0.79)?

12. A bottle weighs 80 grams empty, 280 grams when filled with water, and 250 grams when filled with a medicine. What is the specific gravity of the medicine?

13. An empty bottle weighs 50 grams; the same bottle full of water weighs 200 grams. Some sand is put into the empty bottle and it then weighs 320 grams. Finally the bottle is filled with water, and the bottle, sand, and water weigh 370 grams.

(*a*) Find the capacity of the bottle.

(*b*) Find the volume of the sand.

(*c*) Find the specific gravity of the sand.

14. If one buys 10 pounds of mercury (specific gravity 13.6), how many cubic inches should one get?

15. If the inside of an ice chest measures 24 × 18 × 12 inches, how many pounds of ice (specific gravity 0.92) will it hold?

16. How many pounds of sulphuric acid (specific gravity 1.84) does a 5-gallon carboy contain?

73. City waterworks. Every city has to face the problem of providing a plentiful supply of pure water for household use, for industrial purposes, and for fire protection. Not only must there be enough water, but it must be furnished at sufficient pressure to force it to the tops of high buildings. If the city is located near the mountains, as are Denver and Los Angeles, it is an easy matter to conduct the water from an elevated reservoir in large pipes or **mains** to the houses. Since the water tends to seek its own level, it will rise in the buildings to the height of the reservoir. But in most cities, such as New York, Philadelphia, and Boston, the *gravity system* of waterworks is impossible and a *pumping* system must be employed. The operation of the big steam pumps that are used will be explained later (section 100).

74. Hydrants and faucets. The only parts of this great system of water pipes which we ordinarily see are the **hydrants** on the edge of our sidewalks, and the **taps** or **faucets** at our sinks and bathtubs. These are merely valves for opening and closing the pipes. The internal construction of the ordinary tap is shown in figure 66. The handle operates a screw which forces a disk, faced with a fiber washer, against

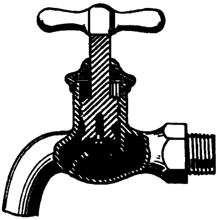


FIG. 66.—Cross section of common faucet.

a circular opening or seat, and so shuts off the water. If the handle is turned the other way, the disk is raised, leaving an opening. This sort of valve may get out of order in two ways: the fiber washer may wear out and the packing about the handle rod may get loose. Both of these can be easily replaced. The packing consists of cotton twine wrapped around the valve stem, and is held in place by what is called a gland.

75. How we measure water pressure. Doubtless we have all found that water flows slowly from a faucet on an upper floor. This is because the water pressure is low there. To measure it, we use some form of **pressure gauge**, and for as small a pressure as this would be, an **open mercury manometer** would be the most accurate form of pressure gauge. It consists of a U-shaped tube filled with mercury, as shown in figure 67.

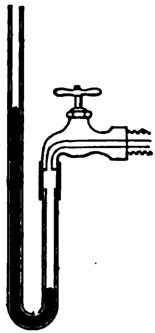


FIG. 67.—Mercury pressure gauge.

Suppose the water pressure is enough to balance a mercury column 4 feet high. How much is the pressure in pounds per square inch? A column of mercury 4 feet high and 1 square inch at the base would contain 48 cubic inches, and would weigh 23.5 pounds. Therefore the pressure of the water would be 23.5 pounds per square inch. With such a gauge it is easy to show that the water pressure is less on the top floor than in the basement.

A mercury gauge is so cumbersome and expensive that a Bourdon spring gauge is generally used. It consists of a brass tube of elliptical section, bent into a nearly complete ring, and closed at one end, as shown in figure 68. The flatter sides of the tube form the inner and outer sides of the ring. The open end of the tube is connected with the pipe

through which the liquid under pressure is admitted. The closed end of the tube is free to move. As the pressure increases the tube tends to straighten out, moving a pointer to which it is connected by levers and small chains. These spring gauges have the scale so graduated that they read directly in pounds per square inch.

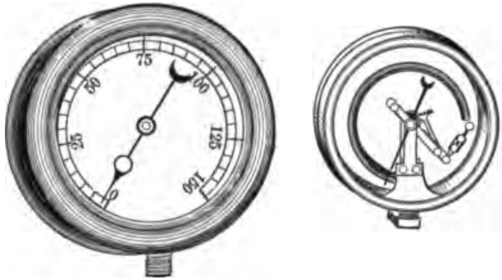


Fig. 68. — Bourdon gauge.

76. Fluctuations in water pressure. Not only does one find a decrease of water pressure in going from the basement to the attic of a house, but if the gauge is attached at one point and watched closely, it will be seen to fluctuate according as much or little water is being drawn elsewhere in the building.

The following experiment shows the same thing on a smaller scale. The tank or reservoir *R* in figure 69 is connected with a supply pipe *AB*. The pressure along the pipe is indicated by the height of the water in the tubes *C*, *D*, and *E*. When the pipe is closed at *B*, the level is the same in *R*, *C*, *D*, and *E*; this is called the static condition. But when the stopper is removed from *B*, and water flows out, the pressure is no longer the same at all points along the pipe, but falls off as the distance from the reservoir *R* increases. This drop in pressure is due to friction against the walls of the pipe through which the water has to run.

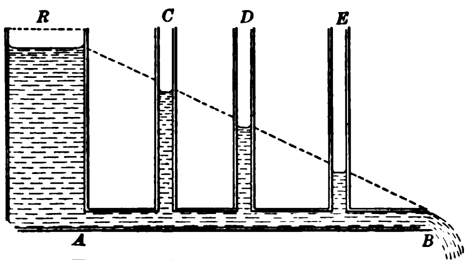


Fig. 69. — Pressure falls with flow.

From this experiment we see that, when a number of faucets are open and the water is flowing, the pressure in the

neighborhood becomes small. To equalize these changes in water pressure and also to provide some flexibility in the system, it is quite common to have a standpipe in the water system nearer the houses than the main reservoir. This also serves as an auxiliary reservoir in case of emergency.

77. Water meter. It is common now to measure the quantity of water which is used by each house. This is done by an instrument called a water meter. There are several types of these meters; one of the simplest is shown in figure 70, and its action is shown in the four diagrams in figure 71. The water enters through the left-hand kidney-shaped opening, shown by dotted lines, and leaves through the similar opening at the right. The moving part, shown by the heavy line, has a hub (the black circle) that travels around in the little circular track provided for it, the moving part meanwhile oscillating to and fro without turning completely around. This makes the various annular spaces enlarge as long as they are in connection with the inlet (watch,

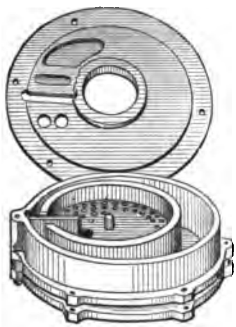


FIG. 70. — Water meter.

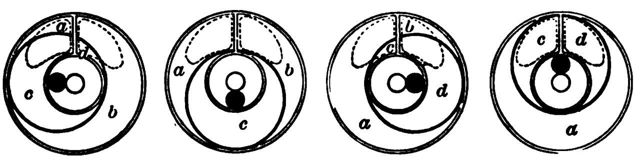


FIG. 71. — Diagrams to show operation of water meter.

for example, the space marked *a* in the diagrams) and contract when they are in connection with the outlet (watch, for example, the space marked *b*). In this way each space measures out its appropriate quantity of water and delivers it to the outlet pipe. The number of revolutions of the hub

is registered on a series of dials (see figure 72) which indicate the number of cubic feet that have passed through the meter. Thus the dials in figure 72 indicate 94,450 cubic feet. The official of the water department reads these dials periodically, and by subtracting can easily compute the water consumed during the period, and so fix the charge in proportion.

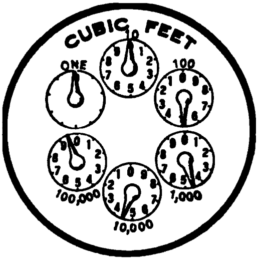


FIG. 72. — Meter dials.

78. Water motors. In cities where water is supplied under considerable pressure, it may be used to run sewing machines, small polishers, grinders, lathes, etc., by means of a water motor. A simple form is shown in figure 73. The stream of water is made to pass through a small opening at high velocity and to strike against some blades or buckets on the rim of a wheel. The wheel is inclosed in a metal case, from which the water flows away to the drain-pipe. The impact of the water against the blades turns the shaft, to which the machines to be driven are connected either directly or by belts.

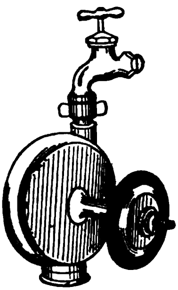


FIG. 73. — Water motor.

79. Water wheels. Just as a small stream of water may be used to turn small machinery, so it is possible to make large streams of water turn large machines, which saw wood, grind corn, and furnish electric lights for streets and houses. Any community possessing a waterfall or a rapid in a river has a valuable source of power. The older types of water wheels were the **overshot**, where the weight of the water slowly turns the wheel, and the **undershot**, where the wheel is let down into a swiftly flowing current. The modern forms of water wheels are the Pelton wheel and the turbine. The little water motor described above is a typical form of **Pelton**

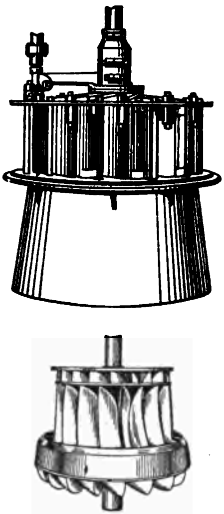


Fig. 74. — Turbine water wheel case and runner.

wheel, the parts of a commercial wheel being the same, but much larger (see plate facing page 72). The efficiency of this type of wheel is much greater than that of the undershot wheel, and sometimes runs as high as 83 %. By far the most important type of water wheel to-day is the **turbine.** This is somewhat like a windmill. The water is conducted from the reservoir above the dam through a cylindrical tube to a "penstock" which surrounds the case of the wheel (Fig. 74). This case stands on the floor of the penstock and is submerged in water to a depth equal to the "head" or height of water supply. The water is not let into the case about the wheel at *one* opening, but through *many inlets* or passages, which are so curved as to direct the water against the blades of the wheel in the most favorable direction to produce rotation, as shown in figure 75. The wheel is attached to a shaft which transmits the power to the machinery above. A small shaft controls the size of the inlet openings in the case. When the water has done its work, it falls from the bottom of the wheel case into the "tail race" below the penstock. These turbines sometimes have an efficiency of 90 %.

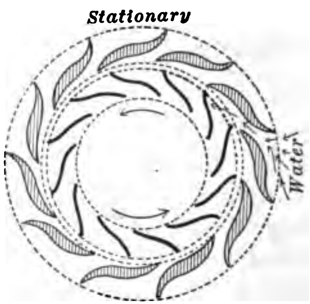


Fig. 75. — Stationary and moving blades of water turbine.

Pelton wheels are used in general where the fall is high and the quantity of water small, turbines where the fall is low and the quantity of water great.

Pelton water wheel, to be installed at Rjukan in Norway, where it will develop 7500 horse power. The wheels of small water motors are similar in shape.

PROBLEMS

1. The water level in a tank on top of a building is 100 feet above the ground. What is the pressure in pounds per square inch at a faucet 10 feet above the ground?

2. If 200 cubic feet of water flow each second over a dam 25 feet high, what is the available power?

3. If the efficiency of the water wheel used at the dam described in problem 2 is 65 %, how many horse power can it supply?

4. How many cubic feet of water must be supplied every second to an overshot wheel which is 20 feet in diameter and delivers 40 horse power at an efficiency of 85 %?

5. The Niagara turbine pits are 136 feet deep, and the average horse power of the turbines is 5000. Their efficiency is 85 %. How many cubic feet of water does each turbine handle per minute?

80. Molecular attractions. When a drop of water falls through the air, it draws itself into an almost perfect sphere. Similarly, lead shot are made by letting molten lead fall from a sieve at the top of a tower into a pool of water at the bottom. In general, a liquid when left to itself tends to get into the shape which has the smallest possible surface, as if it were composed of little particles which had great attraction for one another. It is also observed that there is a great attraction between many pairs of substances if they are brought very close together, as between wood and glue, stone and cement, paint and wood. When this attraction is between particles of the same kind, it is called **cohesion**, and when between particles of different kinds, it is called **adhesion**.

In soap bubbles, it is the cohesion of the little particles of soapsuds which makes the thin film act like an elastic membrane. It is this same property of liquids which makes it possible to lay a somewhat greasy needle on the surface of water and have it float, although steel is eight times as dense as water.

81. Capillarity. Suppose we have two U-tubes (Fig. 76) with their side tubes 30 mm. and 1 mm. in diameter. If we pour water colored with ink into the first tube, and mercury into the second tube. we observe

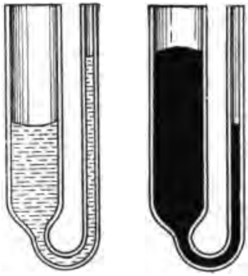


FIG. 76. — Capillarity in small tubes.

that in each case the surfaces in the two sides of the U-tube are not at the same level.

The water wets the surface of the glass and is attracted by it, *i.e.* the adhesion is great. The mercury does not wet the glass, and the cohesion of particles of mercury for each other makes it appear as if there were repulsion between glass and mercury. The surface of the mercury is convex, and it stands at a lower level in the narrow tube than in the wide one. In the case of water, each tube is drawing the liquid up into itself against the pull of gravity. The narrower the tube, the higher the liquid is raised. Since these small tubes have hairlike dimensions, they are called capillary tubes (from Latin *capillus*, a hair), and this phenomenon is called **capillarity**. In this way liquids rise in wicks, in filter paper, and in the soil.

SUMMARY OF PRINCIPLES IN CHAPTER III

Pressure $= \frac{\textbf{force}}{\textbf{area}}$. **Force = pressure × area.**

For liquids under pressure (weight of liquid negligible in comparison): —

Pressure everywhere the same.
Force varies as area.

For liquids with a free surface (weight of liquid the only thing that counts): —

Pressure proportional to depth, independent of direction,
Proportional to density of liquid,
Equal to weight of a column of liquid with a base one unit square and a height equal to the depth.
Average pressure on a surface = pressure at center of surface.
Total force on a surface = average pressure × area.

Archimedes' principle:—

The loss of weight of a body either partly or wholly submerged in a liquid is equal to the weight of the displaced liquid.

If the body just floats, this loss of weight is also equal to the weight of the body.

$$\textbf{Density} = \frac{\textbf{weight of body}}{\textbf{volume of body}}.$$

$$\textbf{Specific gravity} = \frac{\textbf{weight of body}}{\textbf{weight of equal volume of water}}.$$

In the metric system, since 1 cu. cm of water weighs 1 gram,
Density (g. per cm.^{3}) = (numerically) specific gravity.

In the English system, since 1 cu. ft. of water weighs 62.4 lbs.,
Density (lb. per ft.^{3}) = (numerically) 62.4 $\times$ specific gravity.

To get specific gravity:—

Find weight of body.
Find weight of equal volume of water.
Divide.

To get weight of equal volume of water:—

1. **Compute volume. Weight of water = volume $\times$ density of water.**
2. **Loss of weight of body when wholly submerged = weight of equal volume of water. (May have to use sinker.)**
3. **Weigh equal volumes of liquid and of water in a bottle.**
4. **Find loss of weight of a solid in the liquid and in water. (May use either sinker or float, *i.e.* hydrometer.)**
5. **Use balancing columns (see laboratory manual).**

QUESTIONS

1. What advantages has the hydraulic press in testing steam boilers?

2. What device is used to prevent the oil or water from leaking out around the pistons of a hydraulic press?

3. How is Archimedes supposed to have done his famous experiment with the crown?

4. When a ship passes from a river, where the water is fresh, into the ocean, does it rise or sink in the water?

5. If you have a table of densities in the metric system, how could you make a table of specific gravities?

6. How could you determine the specific gravity of a solid soluble in water, but insoluble in kerosene?

7. What is the water pressure in your laboratory?

8. Why has skimmed milk a greater density than normal milk?

9. Two faucets in a town show the same pressure on the gauge and are the same size. If one is one mile from the reservoir, and the other is two miles away, will each faucet deliver the same quantity of water per minute when opened wide?

10. Sometimes when a faucet is opened, especially on an upper floor, the water comes with a rush at first and much more slowly after it has been running a few seconds. Explain.

11. Why does one need to take temperature into account in using the lactometer?

12. What metals float in mercury?

13. How can one pour a liquid out of a glass with the aid of a spoon or glass rod, so that it will not run down the side of the glass?

14. Explain the action of a towel; of a sponge.

15. Explain the process of "fire-polishing" the broken end of a glass tube.

16. If you know the displacement of a battleship, how could you find its weight?

17. Why does one use snowshoes in walking over deep snow?

18. Why is it easier to float when swimming in the ocean than in a river?

19. Why should life preservers be filled with cork instead of hay?

20. A schoolboy in Holland is said to have saved his country from a flood by thrusting his arm into a hole in the dike 150 centimeters below the surface of the sea. Could a small boy hold back the whole North Sea?

CHAPTER IV

MECHANICS OF GASES

Liquids and gases differ in compressibility — air compressors — uses — Boyle's law — vacuum pumps — uses — weight of the air — atmospheric pressure — measured by Torricelli's experiment — barometer and its uses — pressure gauges — lifting effect of air — uses in balloons and pumps for liquids — other properties of gases — absorption and diffusion — molecular theory.

Compressed Air

82. Pneumatic machines. Just as hydraulic machines make use of the properties of liquids, so **pneumatic** machines make use of the properties of **gases**. Nowadays we often clean our houses with vacuum pumps; we stop our express trains with air brakes; we drive the drills and hammers in our shops with compressed air; and we have begun to travel through the air with dirigible balloons and flying machines. To understand the operations of all these machines, we must study the properties of gases.

83. Liquids and gases alike in some respects. Liquids and gases are called **fluids**, because they have no definite shape, but adapt themselves to the shape of the vessel containing them. A liquid, however, has a definite volume under ordinary conditions, filling the lower part of a containing vessel, and being bounded by a free surface above. A gas, on the other hand, has no fixed volume and no free surface, but fills the whole of its containing vessel at once if the vessel is closed, and escapes if the vessel is open at the top. So the little particles of a gas have much more mobility than those of a liquid.

Gases and liquids are alike in that each, when under pressure, distributes that pressure undiminished in all directions in accordance with the principle of Pascal.

The gauges which are used to measure gas pressure are often the same gauges that would be used to measure the pressure exerted by a liquid (see section 75).

84. Air is very compressible. In one respect gases are very different from liquids, namely, in compressibility. This striking difference can be shown in the following experiment.

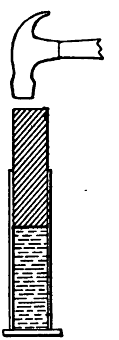


FIG. 77. — Compressibility of fluids.

When a brass tube, with a closely fitting steel rod (Fig. 77), is filled with air, the steel plunger can be easily pushed down by hand, and when the plunger is released, it springs back nearly to its initial position. If it does not come quite back to its initial position, it means that some of the air has leaked out. The entrapped air acts like a spring. But when the tube is filled with water, or any other liquid, it is quite impossible to push the plunger down, to any perceptible extent, by hand, and when the end of the plunger is struck with a hammer, the effect is as if the entire tube were a solid steel column, because the liquid is so nearly incompressible.

85. Air compressors. The simplest form of air compressor is the ordinary bicycle pump, such as is used to inflate the tires on bicycles and automobiles. Figure 78 shows a sectional view of such a pump attached to a tire. It consists of a cylinder C and piston P. On the down stroke some air is entrapped below the piston and compressed, its pressure rising until it becomes equal to that of the air already in the tire. Then the valve S opens, and

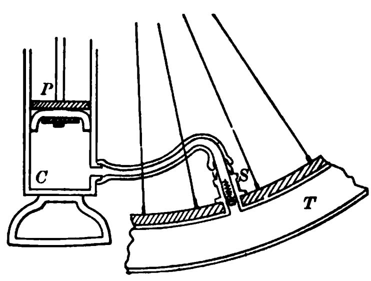


FIG. 78. — Air compressor.

during the rest of the down stroke the air is forced into the tire. When the up stroke starts, this valve closes and the leather washer on the piston bends down and allows air to flow past the piston into the cylinder below. Then on the next down stroke this air, entrapped by the spreading of the leather flange, is compressed and forced over into the tire. There is only one valve, and that is in the stem of the tire.

Large air compressors driven by steam engines or electric motors are much used in steel plants, shops, and quarries to furnish a supply of compressed air. This is delivered as a forced draft to blast furnaces, or stored in steel tanks and used to drive all sorts of pneumatic machinery.

86. Uses of compressed air. There are many tools which are driven by compressed air, such as **riveting hammers** for forming the riveted heads on steel work, and the **pneumatic tools** used in stone cutting, iron chipping, drilling, etc. These are in general lighter and simpler than other portable tools, and there is less danger of fire. When such tools are used in mines, the waste air which they discharge helps to furnish ventilation, and this is often an important advantage. **Rock drills**, and **sand blasts** for cleaning metal and stone surfaces, are other common applications. But perhaps the most interesting application is the **air brake**.

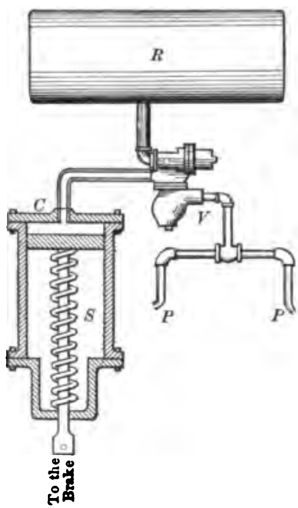


FIG. 79. — Westinghouse air brake.

The essential parts of the Westinghouse air brake are shown in figure 79. *P* is the train pipe leading from a large reservoir on the engine, in which the air is maintained at a pressure of about 75 pounds per square inch. As long as this

pressure is applied to the automatic valve V there is maintained a communication between P and an auxiliary tank R under each car, and at the same time air is cut off from the brake cylinder C. But whenever the pressure in P drops, either by the moving of a lever in the engine cab or by the accidental parting of a hose coupling, the valve V shuts off P and connects the reservoir R with the cylinder C. This pressure on the piston in C forces the brakes against the wheels. As soon as the pressure in the pipe is restored, the valve V reëstablishes the connection between P and R, and at the same time the air in C escapes. The spring S then releases the brakes by pushing up the piston.

87. How volume of air changes with pressure — Boyle's law. In studying about compressed air we are soon confronted with the question as to how much the volume of a given quantity of air changes as the pressure changes. This was first investigated for the case where the temperature of the air does not change during compression, by an Irishman, Robert Boyle (1626–1691), and a few years later by a Frenchman, Mariotte. The results of their experiments showed that if we start with a given volume of air V, subjected to a certain pressure P (Fig. 80), and double the pressure, the volume of air will be reduced to one half. If the pressure be made three times as great, the volume of the air will be reduced to one third, provided the temperature of the air is kept constant. This principle is known as **Boyle's law** and applies to all gases. It may be stated as follows: *The volume of a gas at constant temperature varies inversely as the pressure.*

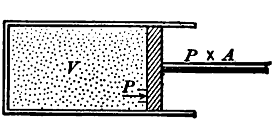


FIG. 80. — Compression of gas.

This may also be expressed in symbols as

$$P : P' :: V' : V \text{ (notice the \textit{inverse} proportion)},$$

or

$$PV = P'V',$$

where P and P' are the pressures, and V and V' the corresponding volumes of a given quantity of gas, kept at some fixed temperature.

At very low temperatures or at very high pressures, this law of Boyle and Mariotte does not hold exactly.

It should be noticed, however, that the air in a bicycle pump does not stay at the same temperature when compressed rapidly, but becomes considerably warmer. The effect of this heating will be discussed in Chapter X. It will then appear that when the air is allowed to get hot, more work has to be done on the pump to produce the same useful result on the tire. The same is true of large compressors, and so it is customary to keep the air in them as cool as possible during compression by circulating water through a jacket around the cylinder, or by spraying water into the cylinder. The ideal case is compression at constant temperature, in accordance with Boyle's law, and large compressors should come as near to this as is practicable.

Problems

NOTE. — Assume constant temperature in these problems.

1. One hundred cubic feet of air under a pressure of 15 pounds per square inch is compressed to 300 pounds per square inch. What does the volume become?

2. The volume of a tank is 2 cubic feet, and it is filled with compressed air until the pressure is 2000 pounds per square inch. How many cubic feet of air under a normal pressure of 15 pounds per square inch were forced into the tank?

3. What is the total force applied to a brake piston 10 inches in diameter, when the pressure is 80 pounds per square inch?

4. One hundred cubic feet of air at a pressure of 15 pounds per square inch are compressed to 36 cubic feet. What is the pressure then?

5. Oxygen is sold in steel tanks under a pressure of 150 pounds per square inch. As the gas is used, the pressure drops. When it has dropped to 50 pounds, what fractional part of the original gas remains? Give your reasoning.

88. Vacuum pumps. We have seen how a bicycle pump can be used to force more air into a given space, and now we shall see how a slight change in the valves will enable us to

suck the air out of a vessel. The first of these so-called "air pumps" was made as long ago as 1650 by a German. Otto von Guericke, then mayor of Magdeburg, who performed numerous experiments. For example, he found that a clock in a vacuum cannot be heard to strike; a flame dies out in it; a bird opens its bill wide, gasps for air, and dies; fish perish; and yet grapes can be preserved six months *in vacuo*. Vacuum pumps used to be found only in physical laboratories, but now they are used so extensively in vacuum cleaners, and in making incandescent lamps and X-ray bulbs, that they are of great commercial importance.

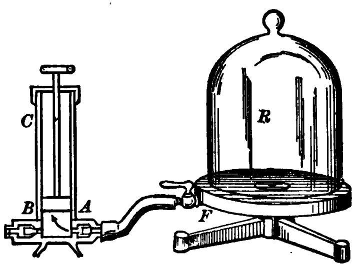


FIG. 81.— Vacuum pump.

A simple form of mechanical vacuum pump is shown in figure 81. It consists of a metal cylinder C fitted with a piston, and having at the lower end two short tubes, A and B, within which are self-acting conical valves, so arranged that the air enters at A and leaves through B.

When the piston is raised, the air in the vessel R, which is to be exhausted, expands into the cylinder C through the valve A. When the piston is pushed down, it compresses this air, closing the valve A and opening the outlet valve B. Thus with each double stroke a certain fraction of the air in the vessel R is removed. It will be seen that even with a mechanically perfect pump we never take out quite all the air; for by each stroke we remove only a *certain fraction* of the air, and the remainder expands to fill the vessel. In practice, no pump is perfect because of leakage. To reduce this, it is common, in "high vacuum" pumps, to cover the piston and the valves with oil, and in some forms made of glass the piston is replaced by a column of mercury.

89. Applications of the vacuum pump. **Vacuum cleaning** (Fig. 82) is an application of the force of suction, created by a vacuum, to the cleansing of buildings and their furnishings. Some vacuum cleaners are portable and some are stationary, some are operated by hand and some by an electric motor, some tend to produce a vacuum by a pump and some by a rotating fan, but the general principle is the same in all.

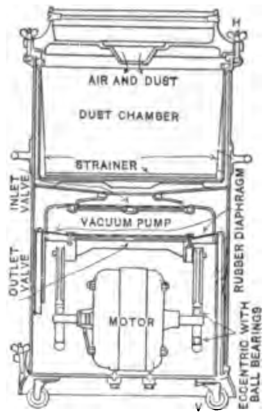


FIG. 82.—Vacuum cleaner.

The problem of getting the air out of **incandescent lamp bulbs** is quite different from that of vacuum cleaning, in that we have a very limited space to be exhausted and this must be very completely pumped out. Usually less than one millionth part of the air is left in the bulb. For this purpose two mechanical pumps are worked in tandem, one to take air directly from the bulb, and the other to take it from the cylinder of the first pump. After these have done their work, the air remaining is still further reduced by burning phosphorus or some other combustible in the bulb. **X-ray tubes** are made in much the same way, except that the process must be continued longer so as to produce a still more rarified condition of the air.

WEIGHT OF THE AIR

90. Density of air. We are so accustomed to having air about us that we do not ordinarily think of it as having either *volume* or *weight.* We speak of an "empty" bottle when we usually mean a bottle filled with air. Yet when we try to fill a narrow-necked bottle with a liquid, we find that we can make the liquid run in only as fast as the air gets out. If we push a glass tumbler mouth down into a

pail of water, it is not filled with water, because it is filled with air. Air occupies space just as does any other fluid.

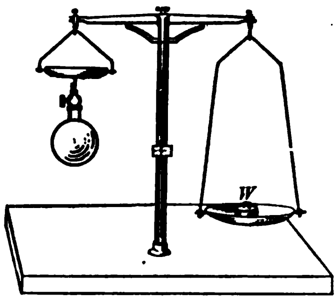


FIG. 83. — Proof that air has weight.

Furthermore, air and other gases have *weight*, although we seldom realize this fact.

In order to make it evident that air has weight, let us try the following experiments. Suppose we carefully counterbalance on the scales (Fig. 83) a hollow metal vessel (a tin can with a bicycle valve soldered into the top will serve). Then if we pump more air into the vessel and put it on the scales again, we find that it has gained in weight. If we repeat the process, we find that it weighs a little more after each pumping.

In the same way if we take a vessel from which the air has already been exhausted, such as an electric light bulb, carefully counterbalance it, and then let the air in by filing off the tip, we find that the scalepan containing the bulb and the broken pieces goes down, showing that the air which has entered has weight.

Careful experiments show that under ordinary conditions a liter of air weighs about 1.3 grams, or 1 cubic foot weighs about 1.3 ounces.

Since gases have both **volume** and **weight**, we may express their **densities** in the usual way, as so many grams per cubic centimeter or so many pounds per cubic foot. For example, the density of ordinary air is about 0.0013 grams per cubic centimeter, or 0.08 pounds per cubic foot. Many gases have densities even smaller than that of air. Thus the density of hydrogen under standard conditions is only 0.000090 grams per cubic centimeter.

Evidently if the pressure on a certain volume of air is doubled, the volume is halved, and the air becomes twice as dense. In other words, *the density of air or of any gas varies directly as the pressure at constant temperature.*

Problems

1. If a liter of air weighs 1.3 grams, how much does the air in a room weigh, if the room is 3 meters high, 10 meters long, and 8 meters wide?

2. If the pressure in a compressed-air tank is 150 pounds per square inch, what does 1 cubic foot of this compressed air weigh? (1 cubic foot under a pressure of 15 pounds weighs 1.3 ounces.)

3. A spherical balloon 10 meters in diameter is filled with hydrogen. Find the weight of the hydrogen.

4. A bicycle tire has about the same volume as a cylinder 85 inches long and 1 inch in diameter. If you pump your tires up from 15 pounds to 75 pounds per square inch, how much more will the bicycle weigh than before?

5. A compression pump, whose capacity is 500 cubic centimeters, is used to force air into a can whose volume is 1 liter. What is the density of the air after 3 complete strokes?

6. A vacuum pump, whose capacity is 500 cubic centimeters, is used to exhaust the air from a liter flask. What is the density of the air left in the flask after 3 complete strokes?

91. Pressure of the atmosphere. Since we are living at the bottom of an ocean of air, and this air is a fluid which has weight, it is natural to expect that it exerts a pressure. Ordinarily we are not aware of this pressure because it pushes up on the bottoms of objects almost as much as it pushes down on the tops of them. If we could get rid of this upward pressure underneath, we would see how great the downward pressure on top really is. This can be done with a vacuum pump, or in part even with the lungs.

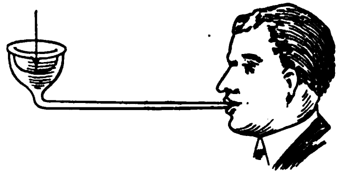


Fig. 84.—Removing the upward pressure of the air.

Let us fasten a piece of sheet rubber over the end of a thistle tube, as shown in figure 84. If we suck the air out of the bulb with the mouth, the rubber is forced downward because of the atmospheric pressure.

This experiment is even more striking when performed with a larger membrane and with a vacuum pump. If we tie a piece of rubber over the mouth of the glass vessel shown in

figure 85, and gradually pump out the air, the rubber will be pushed down more and more by the pressure of the air above it, until it finally bursts. If a piece of bladder is used instead of rubber, it will break with a loud report.

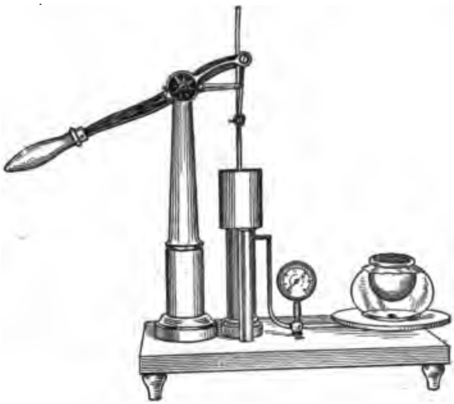


FIG. 85. — Air pressure breaking a membrane.

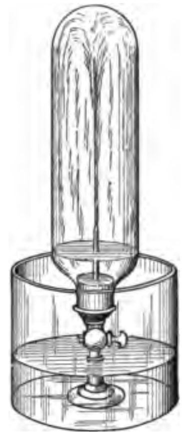
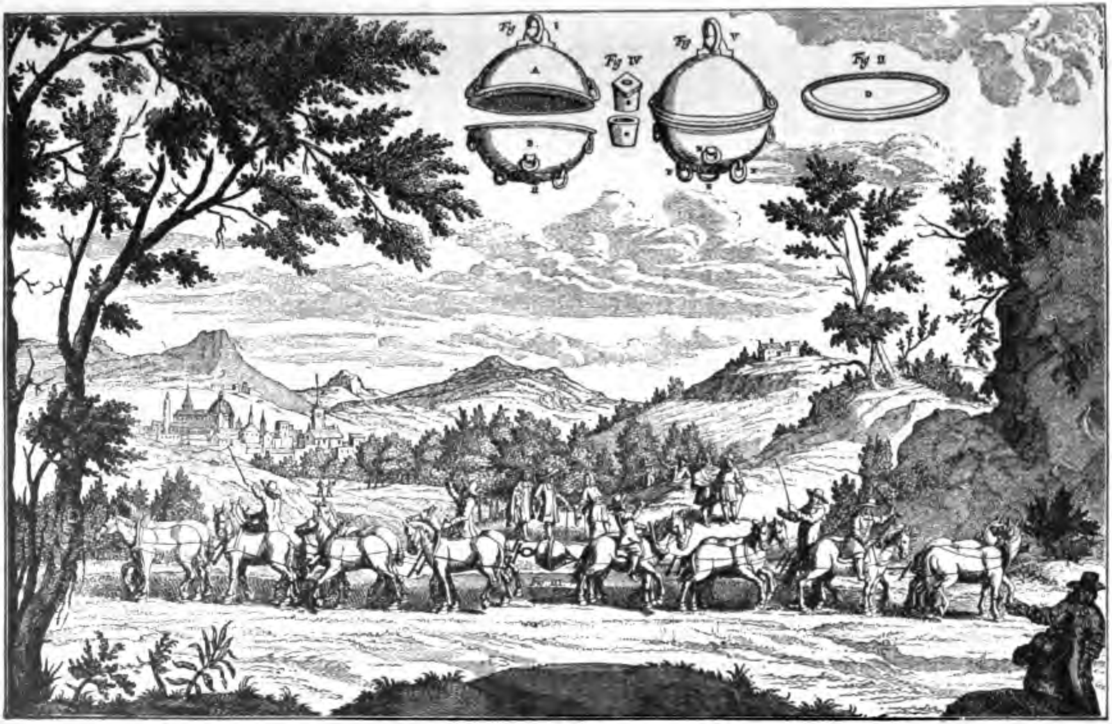


FIG. 86. — Fountain *in vacuo*.

If we pump the air out of a tall glass vessel provided with a stopcock and jet tube, and then place the mouth of the jet tube under water and open the stopcock, we see the water rushing up into the vacuum like a fountain. How can we determine how much air was removed?

One of the most interesting of Otto von Guericke's experiments was that with his famous "Magdeburg hemispheres." These were two hollow hemispheres a little over a foot in diameter which fitted together so well that the air could be pumped out from between them. The pressure of the surrounding atmosphere then held them together firmly. In a test before the Reichstag and the Emperor, it required sixteen horses, four pairs on each hemisphere, to pull them apart.

92. "Nature abhors a vacuum." The ancients tried to explain many phenomena by saying that "nature abhors a vacuum," but when the great Italian philosopher, Galileo (1564–1642), found that a suction pump would not raise



The experiment with the Magdeburg hemispheres, performed by Otto von Guericke.

water more than 33 feet, he remarked that nature's horror of a vacuum was a curious emotion if it stopped suddenly at 33 feet. He already knew both that air has weight and that the "resistance to a vacuum" was measured by a column of water about 33 feet high, yet he left it to his friend and successor, Torricelli (1608–1647), to unite these two ideas.

93. Torricelli's experiment. Torricelli devised a means of measuring this "resistance" which nature "offers to a vacuum" by a column of mercury in a glass tube instead of a column of water.

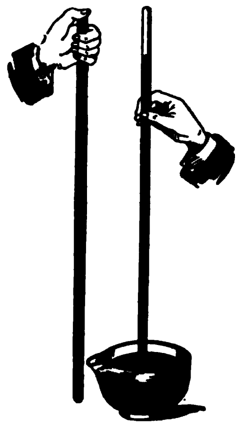


FIG. 87.—Torricelli's experiment.

We may repeat this experiment if we take a stout glass tube about 3 feet long, closed at one end, and fill it completely with mercury. If we close the opening with the finger, invert the tube, and put its open end into a tumbler of mercury, we observe that, when the finger is removed, the mercury in the tube (Fig. 87) sinks to a level about 30 inches above the mercury surface in the tumbler. If we incline the tube to one side, the metal fills the entire tube and hits the top of the glass with a sharp click. The space above the mercury is empty except for a minute quantity of mercury vapor. It is, indeed, the most perfect vacuum that we know how to make.

The column of mercury in the tube just balances the pressure of the atmosphere on the mercury in the larger vessel at the bottom. In other words, liquids rise in exhausted tubes because of the pressure exerted by the atmosphere on the surface of the liquid outside, and not because of any mysterious sucking power created by the vacuum.

94. How to calculate the pressure of the atmosphere. From the law that pressure in a heavy liquid is everywhere the same at the same depth, we know that the pressure on the mercury in the dish (Fig. 88) is the same at *a* as outside.

Outside this pressure is exerted by the atmosphere. At *a* it is exerted by the column of mercury *ab*. Under standard conditions, the pressure at *a*, that is, the force per square centimeter, is evidently equal to the weight of a column of mercury 76 centimeters high and 1 square centimeter in cross section. This is the weight of 76 cubic centimeters of mercury, or 76 times 13.6 grams, or 1034 grams.

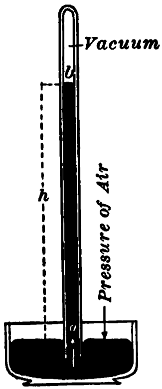


FIG. 88.—Mercury column supported by air.

In the English system it is the weight of a column of mercury about 30 inches high and 1 square inch in cross section; that is, 30×0.49, or 14.7 pounds. Roughly, then, *one "atmosphere" is about 1 kilogram per square centimeter or about 15 pounds per square inch.*

95. Pascal's experiments. Pascal reasoned that if the mercury column was held up simply by the pressure of the air, the column ought to be shorter at a high altitude. So he carried a Torricelli tube to the top of a high tower in Paris, and found a slight fall in the height of the mercury column. Desiring more decisive results, he wrote to his brother-in-law to try the experiment on the *Puy de Dôme*, a high mountain in southern France. In an ascent of 1000 meters, the mercury sank about 8 centimeters, which greatly delighted and astonished them both.

Pascal also tried Torricelli's experiment, using red wine and a glass tube 46 feet long, and found that with a lighter liquid a much higher column was sustained by the pressure of the air. These experiments were carried out in 1648, five years after Torricelli's discovery.

96. The barometer. The arrangement constructed by Torricelli may be set up permanently as a means of measuring the pressure of the atmosphere. It is then called a

barometer. To "read the barometer" means simply to measure accurately the height of the mercury column above the surface of the liquid in the reservoir. In the form of barometer shown in figure 89 this reservoir has a flexible bottom which may be raised or lowered so as to bring the surface of the mercury to the zero point of the scale which is at the tip of a point projecting into the reservoir. The height of the mercury is then read by observing the position of the liquid in the tube.

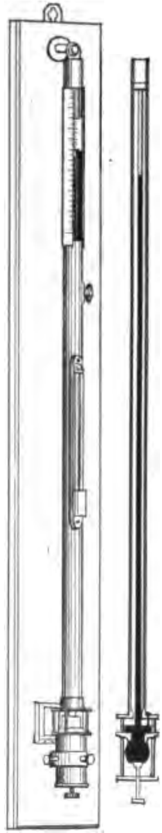


FIG. 89. — Mercurial barometer.

A more convenient form to carry about is the **aneroid** or metallic barometer (Fig. 90). As the name indicates, it is "without liquid" and consists essentially of a disk-shaped metal box, which has a thin corrugated metal top. When the air has been pumped out of the box, it is sealed up, its top being supported by a stout spring to prevent its collapsing. As the pressure of the air changes, the top of the box moves up or down, and the small motion is greatly magnified by means of levers and a delicate chain, and is communicated to a pointer which moves over a scale. A hairspring serves to take up the slack of the chain. The scale is graduated to correspond to the readings of a standard mercurial barometer. Aneroid barometers are made in various sizes. Some are even as small as ordinary watches.

FIG. 90. — Aneroid barometer.

97. Uses of the barometer. The barometer indicates changes in atmospheric pressure. These changes may be due to fluctuations in the atmosphere itself or to changes in the elevation of the observer.

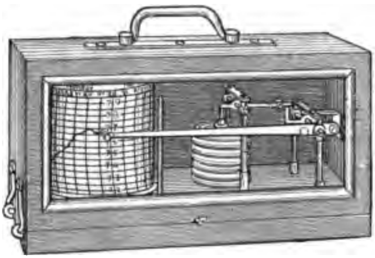


FIG. 91. — Barograph, or self-recording barometer.

If a barometer, kept always at the same elevation, is frequently observed, or if it makes a continuous record, as does a barograph (Fig. 91), it is found to fluctuate according to the **weather.** Experience shows that a "falling barometer," that is, a sudden decrease of atmospheric pressure, precedes a storm; and a "rising barometer," that is, an increasing atmospheric pressure, indicates the approach of fair weather; while a steady "high barometer" means settled fair weather.

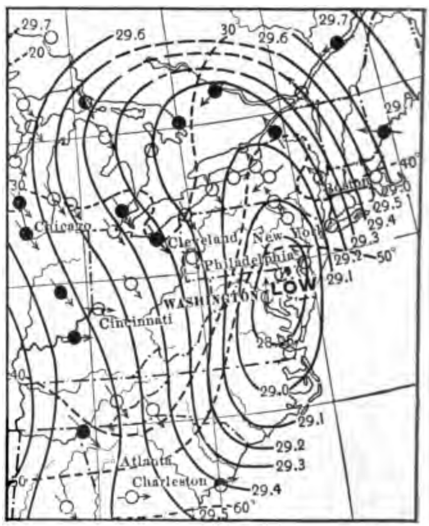


FIG. 92. — Portion of a weather map.

The Weather Bureau has barometric readings taken simultaneously at many different places, and the results are telegraphed to central stations, where weather maps are prepared. On these maps it is observed that there are certain broad areas where the pressure is **low,** and other sections where the pressure is **high.** The areas of

low barometric pressure are usually storm centers, which move in a general easterly direction. If we know where these low pressure areas are located and their probable movement, we may predict the weather. Figure 92 shows a portion of a government weather map. The curved lines, showing the places where the barometric pressure is equal, are called **isobars.** The direction of the wind at places of observation is indicated by an arrow, and it will be noticed that these arrows usually point from a "high" to a "low." A careful study of these phenomena (which is called meteorology) shows that these "lows" are really great eddies of air slowly moving in a counterclockwise direction about the center of the low.

Another important use of the barometer is to measure the **difference in altitude** of two places. If a surveyor or explorer carries a barometer up a mountain, he notices that it indicates a decrease in atmospheric pressure as he ascends. For places not far above sea level this decrease is about 1 millimeter for every 11 meters of elevation or 0.1 of an inch for every 90 feet of ascent. Aneroid barometers graduated in feet or meters are always carried by balloonists and aviators to tell how high they are.

98. Pressure gauges. Besides barometers, which are really pressure gauges designed for pressures of one atmosphere or less, we need gauges for *higher pressures* such as those in a steam boiler, or a compressed-air tank, and gauges for *very low pressures*, such as those in the condenser of a steam engine or a vacuum pump.

To measure slight differences in pressure, the **open manometer,** described in section 75, is used, usually with some liquid lighter than mercury as the indicating fluid.

If we bend a piece of glass tubing as shown in figure 93, and partly fill the tube with colored water, we have a suitable gauge to measure the pressure of ordinary illuminating gas, which will usually cause a difference in levels, *A*, *B*, of about 2 inches.

For high pressures this form of gauge, even when filled with mercury, becomes too cumbersome, so a **closed manometer**, like that shown in figure 94, is used. The mercury stands at the same level in both arms, when the pressure is one atmosphere. If the pressure is greater than this, the mercury is forced into the closed arm, compressing the confined air according to Boyle's law. The scale may be made to read in atmospheres.

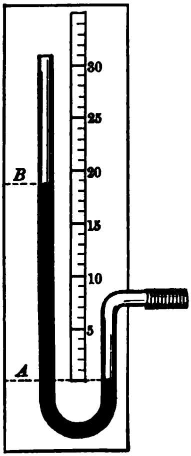


Fig. 93. — Open manometer.

For practical work, the **Bourdon spring gauge**, described in section 75, is used. Such gauges are usually graduated so as to read zero when the pressure is really one atmosphere; that is, they indicate the difference between the given pressure and atmospheric pressure. Therefore when an engineer speaks of a pressure of 100 pounds "**by the gauge**," he means 100 pounds per square inch *above* one atmosphere; when he means the total pressure above a vacuum, he usually says "100 pounds **absolute**."

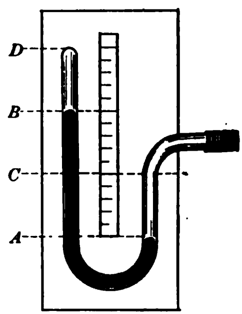


Fig. 94. — Closed manometer.

When pressures less than one atmosphere are to be measured, such as the vacuum in the condenser of a steam engine (section 219), a barometer of the ordinary form would be inconvenient because the whole reservoir or cup at the bottom would have to be exposed to the pressure to be measured. The gauge is, therefore, arranged so as to admit the low pressure to be measured to the top of the barometer tube. The height of the mercury then indicates the *difference* between the small pressure and that of the atmosphere. The better

the vacuum, the higher such a gauge reads. Thus engineers usually speak of a 26 or a 28 inch vacuum, meaning a pressure less than the standard 30 inch atmosphere, by 26 or 28 inches of mercury. The best vacuums now obtained in steam turbine condensers are from 29 to 29.5 inches. Since these mercury gauges would be inconvenient in engine houses, Bourdon gauges are used. They are graduated to read in inches like the mercury gauges which they replace.

PROBLEMS

1. A diver works 51 feet below the surface of the water. To how many atmospheres of pressure is he subjected?

2. When the barometer reads 74.5 centimeters, how many inches does it read?

3. When a mercury barometer reads 76 centimeters, what would a glycerine barometer read? (The density of glycerine is 1.26 grams per cubic centimeter.)

4. When the barometer reads 75 centimeters, what is the atmospheric pressure in grams per square centimeter?

5. During a storm the barometer "dropped" 1.5 inches. How far would a water barometer have fallen?

6. If a certain pressure is 75 pounds per square inch, how many kilograms per square centimeter is it?

7. During a mountain climb the barometer falls 1.75 inches. What is the net height climbed (in feet)?

8. Two glass tubes are arranged vertically (Fig. 95) so that their lower ends dip into water and kerosene, respectively, while their upper ends are joined to a mouthpiece. When some of the air in the tubes is sucked out, the water rises 26 centimeters and the kerosene 33 centimeters. Find the specific gravity of the kerosene. (This is a common way of getting specific gravity.)

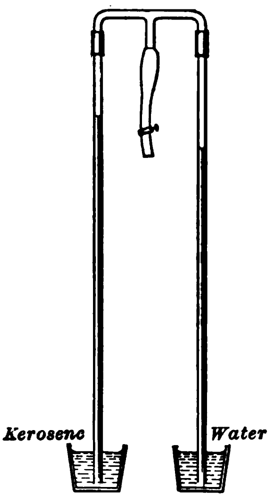


FIG. 95. — Specific gravity by balanced columns.

9. How much force is exerted against an 8-inch piston of an air brake when the pressure is 90 pounds "by the gauge"?

10. The original Magdeburg hemispheres are preserved in a museum at Munich. They are about 1.2 feet in diameter inside. When the air was exhausted, it is said to have required 8 horses on each half to separate them. Assuming that the pressure of the atmosphere was 15 pounds per square inch, find the force exerted by each set of horses. (Reckon pressure on circle 1.2 feet in diameter. Why?)

99. The lifting effect of air. We have seen that when one climbs a mountain, the pressure of the air decreases. A sensitive barometer will indicate this decrease of pressure even when it is lifted from the floor to a table. Therefore the upward pressure of the air on the bottom of any object is slightly more than the downward pressure of the air on the top. In other words, just as in the case of liquids, there is a lifting effect on everything surrounded by air, and this lifting effect is equal to the weight of the air which is displaced.

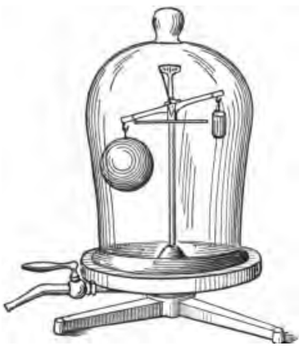


FIG. 96. — Lifting effect of air.

To make this principle of the buoyancy of the air seem more real, let us balance a hollow brass globe against a solid piece of brass under the receiver of a vacuum pump (Fig. 96). When the air is pumped out, the globe seems to be heavier than the solid brass weight, because the support of the air around it has been withdrawn. If the air is readmitted rapidly, the rise of the globe will be very apparent.

Most things are so heavy in comparison with the amount of air they displace that this loss in weight, due to the buoyancy of the air, is not taken into account. For example, a barrel of flour would weigh about 8 ounces more *in vacuo* than in air. But if the volume of air displaced is very large and the weight small, as in the case of a balloon, the object is lifted just as a piece of wood is lifted when immersed in water. A balloon is usually made of cloth which is treated with a special varnish to make it as nearly gas-tight as possible,

and is surrounded by a network of ropes and cords to hold up the car and its load. The bag is filled with hydrogen, which, volume for volume, is only one fourteenth as heavy as air. Sometimes, for short trips, illuminating gas, or even hot air is used. Of course a large part of the lifting force is used in raising the car, the rigging of the balloon, and the silk of which the bag is made. The rest is available for lifting passengers and ballast. To compute the lifting force of a balloon we have only to get the difference between the weight of the air displaced and the weight of the hydrogen, gas, or hot air.

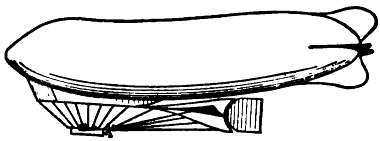


FIG. 97. — A dirigible balloon.

The dirigible balloon (Fig. 97) is provided with propellers driven by gas engines, and rudders to steer with. But the bag has to be made so large to support the weight of all this machinery, that the balloon is much at the mercy of the wind.

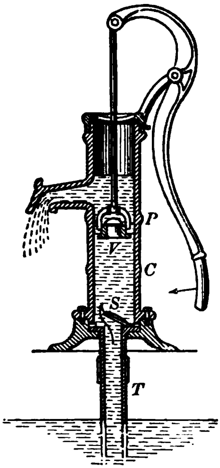


FIG. 98. — A suction pump.

100. Pumps for liquids. The ancients used pumps to lift water from wells, even though they did not know why a pump works; they thought it was because "nature abhors a vacuum." We know now that the underlying principle is the same as in a mercurial barometer: it is the pressure of the atmosphere on the surface of the water in the well that pushes the water up into the pump.

For example, let us consider the ordinary **suction pump** shown in figure 98. This consists of a cylinder *C*, which is connected with the well or cistern by a pipe *T*. At the bottom of the cylinder is a clapper valve *S*, opening up. A piston

P can be worked up and down in the cylinder by means of a handle. This piston also contains a valve opening up. On the ***up stroke*** of the piston *P*, the valve *V* remains closed because of its weight and the pressure of the air upon it. Between the piston and the bottom of the cylinder there would be a partial vacuum if the valve *S* remained closed. But the pressure of the air on the water in the well forces some water up through the pipe *T*, past the valve *S* into the cylinder *C*. On the ***down stroke*** of the piston the valve *S* closes, the valve *V* opens, and the water gets above the piston. On the next up stroke it is lifted out through the spout. The valve *S* must never be more than 34 feet above the water in the well, and in practice this distance is seldom more than 30 feet. Why?

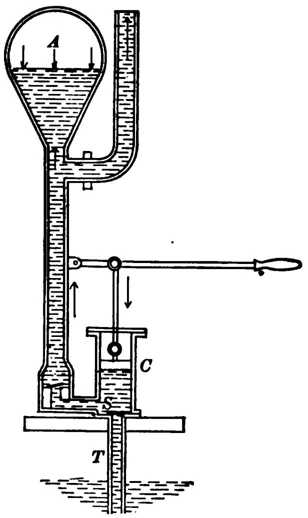


FIG. 99. — A force pump.

Another kind of pump, shown in figure 99, is called a **force pump**. The suction pipe *T* with its valve *S* are exactly like the corresponding parts of the house pump just described, but the piston has no opening through it, and the outlet pipe and a second valve are at the bottom of the cylinder. Raising the piston fills the cylinder with water; pushing it down again forces the water out through the second pipe. If enough force is exerted on the piston, the water can be pushed up to a considerable height. The pump can therefore be located near the bottom of a well or mine shaft.

Since the water is forced up only on the down stroke, it comes in jerks. To reduce the jar and shock, an air chamber

A is connected with the delivery pipe, so that the air may act as a cushion or spring. Power pumps, such as are used on fire engines, or in city waterworks, are "double acting" (Fig. 100), which gives a still steadier stream.

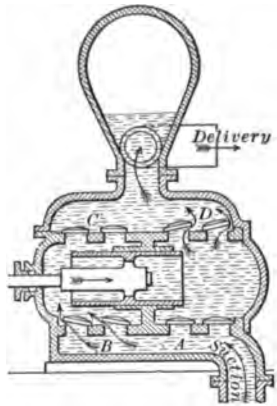


FIG. 100. — A double-acting force pump.

When a large volume of water is to be lifted a short distance, a **centrifugal pump** (Fig. 101) is used. This is something like a water wheel worked backwards. As the wheel inside (Fig. 102) is turned, the water, which enters near the hub, gets caught between the blades and is hurled outward into the delivery space around the wheel, even against some pressure there. Similar machines, called "blowers," are used to force a current of air through a building for ventilation, or to make "forced draft," for furnaces. Often several of these pumps are used in series to give higher pressures. Large turbine pumps of this sort, driven by steam turbines, have recently begun to revolutionize blast furnace practice in the United States, because of the extremely steady rate at which they furnish the air needed for combustion.

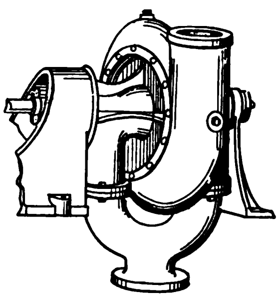


FIG. 101. — Centrifugal pump.

Another form of pump is the "**air-lift**" **pump** (Fig. 103). Its action depends upon the formation of a column of mixed *water and air* which, because of

FIG. 102. — Section of centrifugal pump.

its lesser specific gravity, is raised by a shorter column of water. Such a pump will lift water mixed with air as much as 40 feet above the level of the water. It consists of two tubes, the smaller of which is centered within the larger. The smaller pipe conveys compressed air down into the water to be lifted. The mixture of water and air rises through the outer tube. This sort of pump is cheap to make, is simple in its operation, and has no wearing parts ; but its efficiency is low. It would be especially useful in an artesian or oil well if the water or oil naturally stands too far below the surface to be reached by a suction pump and if the well is so small that a force pump cannot be put down into it.

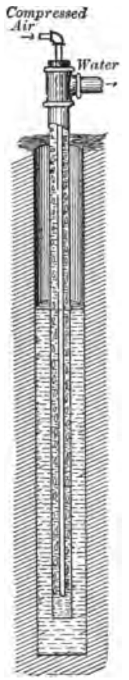


Fig. 103. — Air-lift pump.

101. Siphon. The siphon is a bent tube with unequal arms. It is used to empty bottles and tanks which cannot be overturned, or to draw off the liquid from a vessel without disturbing the sediment at the bottom. If the tube is filled and placed in the position shown in figure 104, the liquid will flow out of the vessel A and be discharged at a lower level D. The force which makes it flow is the weight of the column of water CD, which is between the water level AA' and the water level DD'. If the water level DD' is raised to AA', this moving force becomes nothing and the water ceases to flow ; if the level DD' is lifted above AA', the liquid flows back into the vessel A. A siphon works, then, as long as the free surface of the liquid in one vessel is lower than the free surface of

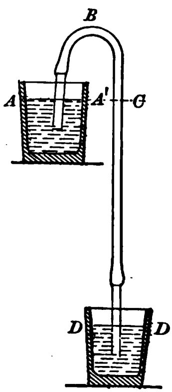


Fig. 104. — A siphon.

the liquid in the other vessel. A water siphon will not work if the top of the bend B is more than 34 feet above the level AA'. Why?

Siphons are often used on a large scale in engineering. For instance, in power plants the water used to condense the steam is often taken from the ocean, raised 10 or 15 feet to the condenser, and carried back to the ocean, through a pipe that is everywhere air-tight and acts like a siphon. The only work that the pumps have to do is to keep the water moving against the friction in the pipe. A large inverted siphon is used at Storm King on the Hudson River, to carry the water supply of New York City under the river, some 700 feet below its surface. The lifting of the water on one side is done by the water descending from a slightly greater height on the other.

Siphons on a smaller scale are used in every aqueduct to carry water over hills or across valleys. In such cases air bubbles carried along in the water tend to collect at the top of every hill, and so small air pumps have to be installed to keep the pipes full of water.

PROBLEMS

1. How many feet could water be lifted with a perfect suction pump (*a*) at sea level, and (*b*) in Denver, Col. (altitude about 5000 ft.)?

2. How many feet could crude oil (density 0.89 grams per cubic centimeter) be lifted out of an oil well by a perfect suction pump at sea level?

3. How much work is needed to lift 100 gallons of water 25 feet with a perfect pump?

4. How much power is needed to raise 100 gallons of water per minute 25 feet with a perfect pump?

5. A force pump is to deliver water at a point 20 feet above the level of its barrel. How great is the water pressure in the barrel when the piston is descending?

6. The piston of a fire-engine force pump is 4 inches in diameter, and the total force exerted on it by the engine is 600 pounds. If the pump acts perfectly, at how great a height will it deliver water?

7. A siphon is to be used to transfer mercury from one bottle to another. How far above the level of the mercury in the higher bottle can the top of the siphon tube be?

Other Less Important Properties of Gases

102. Absorption of gases in liquids. If we slowly heat a beaker containing cold water, small bubbles of air will be seen to collect in great numbers upon the walls (Fig. 105) and to rise through the liquid to the surface. It might seem at first that these are bubbles of steam, but they must be bubbles of air, first because they are formed at a temperature below the boiling point of water, and second because they do not condense as they come to the cooler layers of water above.

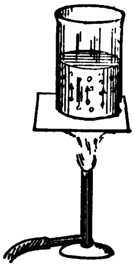


Fig. 105. — Bubbles of air in water.

This simple experiment shows that ordinary water contains dissolved air, and that the amount of air which water can hold decreases as the temperature rises. It is the oxygen of the air that is dissolved in water which supports the life of fish. The amount of gas absorbed by a liquid depends on the pressure of the gas above the liquid. Thus soda water is ordinary water which has been made to absorb large quantities of carbon dioxide gas by pressure. When the pressure is relieved, the gas escapes in bubbles, causing effervescence. Careful experiments show that the amount of gas absorbed is proportional to the pressure. The amount of gas which will be absorbed by water varies greatly with the nature of the gas. For example, at 0° C. and at a gas pressure of 76 centimeters of mercury, 1 cubic centimeter of water will absorb 0.049 cubic centimeters of oxygen, 1.71 cubic centimeters of carbon dioxide, and 1300 cubic centimeters of ammonia gas. The ordinary commercial aqua ammonia is simply ammonia gas dissolved in water.

103. Absorption of gases in solids. Certain porous solids, such as charcoal, meerschaum, silk, etc., have a great capacity for absorbing gases. For example, charcoal will absorb 90 times its volume of ammonia gas and 35 volumes of carbon dioxide. It is this property of charcoal which makes it use-

ful as a deodorizer. This absorption seems to be due to the condensation of a layer of gas on the surface of the body or of the pores within the body. Platinum in a spongy state absorbs hydrogen gas so powerfully that if a small piece is placed in an escaping jet of hydrogen, the heat developed by the condensation is enough to ignite the jet. This has been made use of in self-lighting Welsbach mantles.

A familiar example of the absorption of gases by liquids and solids is the contamination of milk and butter by onions, fish, or other kinds of food, if they are kept in the same compartment of a refrigerator. Onions, for instance, give off a small quantity of gas which we can easily detect by our sense of smell, or by the watering of our eyes. This gas, when absorbed by milk or butter, affects its taste.

104. Diffusion of gases. One of the difficulties in the successful construction of balloons is due to the **diffusion** of the gas through the bag. The diffusion of hydrogen through a porous cup is shown in the following experiment.

If we set up a porous cup with a stopper and glass tube, as shown in figure 106, and allow hydrogen (or illuminating gas) to fill the jar which surrounds the porous cup, we observe bubbles rising from the end of the glass tube, which dips under water. This means that the gas is going through the porous walls of the cup and forcing the air out at the bottom. If we now shut off the gas and remove the jar, we presently see the water slowly rising in the tube, which shows that the gas inside the cup is going out.

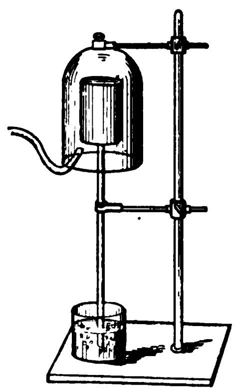


FIG. 106.—Diffusion of hydrogen through porous cup.

The fact that a little ammonia (or any other gas with a powerful odor) introduced into a room is soon perceptible in every part of the room shows that the gas particles travel quickly across the room. Moreover, this mixing of gases goes on whatever

the relative densities of the gases, so that a heavy gas like carbon dioxide and a light gas like hydrogen will not remain in layers like mercury and water, but will quickly diffuse and become a homogeneous mixture. Experiments show that the smaller the density of the gas, the greater the velocity of its diffusion.

105. Molecular theory of gases. To explain the pressure of gases and their diffusion, it is now generally supposed that all substances are made of very minute particles called **molecules.** These molecules are so minute that we cannot see them even with the most powerful microscopes. In one cubic centimeter of a gas there are probably not less than 10^{19} (that is, 1 followed by nineteen ciphers) molecules. The spaces between these molecules are supposed to be much larger than the molecules themselves. This explains why gases are so easily compressed and diffuse so quickly.

Then, too, these little particles are supposed to be flying about in all directions with great velocity. They are supposed to travel in straight lines except when they hit each other and bounce off. Gas molecules seem to have no inherent tendency to stay in one place, as do the molecules of solids. This explains why gases fill the whole interior of a containing vessel. This also explains gas pressures, for the blows which the innumerable molecules of a gas strike against the surrounding walls constitute a continuous force tending to push out these walls. When a gas is compressed to half its volume, the pressure is doubled, because doubling the density doubles the number of blows struck per second against the walls. It has even been possible to calculate the molecular velocity necessary to produce this outward pressure. It appears that the molecules of gases under ordinary conditions are traveling at speeds between 1 and 7 miles per second. The speed of a cannon ball is seldom greater than one half a mile per second.

This, in brief, is the so-called kinetic theory of gases.

SUMMARY OF PRINCIPLES IN CHAPTER IV

Pascal's Law of Transmission of Pressure: For gases under pressure, the *pressure* is everywhere the same; the *force* varies as the area.

Boyle's Law: *Volume* of gas at constant temperature varies *inversely* as *pressure*.

Density of a gas varies *directly* as *pressure*.

Lifting effect of air is equal to *weight of air displaced*.

Atmospheric pressure equal to about
30 inches of mercury,
34 feet of water,
15 pounds per square inch,
1 kilogram per square centimeter.

QUESTIONS

1. Why can Torricelli's experiment be performed as well indoors as outdoors?

2. How and why can a glass of water be inverted with the aid of a card without spilling the water?

3. Why does a rubber tube often collapse when connected with a vacuum pump? Why does not a rubber tube always collapse when connected with a vacuum pump?

4. Why must a mercurial barometer be hung in a vertical position?

5. What would be the result of putting a mercurial barometer under a tall bell glass on an air pump?

6. Would a siphon work in a vacuum?

7. What would be the effect of lengthening the long arm of a siphon?

8. A boat lying on a beach is full of water. How could you empty it with the help of a suitable length of rubber hose? Could you use the same method to get the bilge water out of a boat floating in the water?

9. Why are not barometers filled with water?

10. What advantage has a pneumatic automobile tire over a solid tire of the same size?

11. How can a balloon be made to sink or rise?

12. Why does a man under water in a diving suit have to be supplied with *compressed* air?

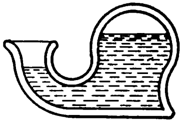


FIG. 107. — Inkwell.

13. Explain why the liquid does not run out of a medicine dropper.

14. Explain the action of a so-called "pneumatic" inkstand (Fig. 107), or of a drinking fountain (Fig. 108), or of a poultry fountain (Fig. 109).

15. A man finds that cider does not flow out of a barrel until he removes the bung. Explain.

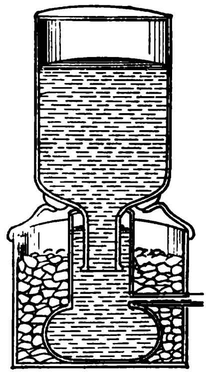


FIG. 108. — Drinking fountain.

16. A vessel 1 meter deep is filled with mercury. Can it be entirely emptied by means of a siphon?

17. Why does a chemist usually reduce the volumes of gases to standard pressure, that is, 76 centimeters?

18. What advantages has compressed air over electricity for the transmission of power?

19. If the area of a man's body is 20 square feet, what is the total force exerted on him by the atmosphere? Why is he not crushed by this force?

20. What facts indicate that the atmosphere becomes rarer and rarer as one rises above sea level?

21. In building tunnels workmen usually have to work in chambers filled with compressed air. Why is this necessary?

22. Get the dimensions and weights of some of the large balloons used in international races, and compute their lifting power. Estimate the amount of ballast that can be carried in addition to the weight of the balloon, car, and passengers.

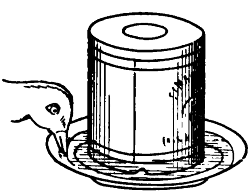


FIG. 109. — Poultry fountain.

23. How does a gas meter work?

24. Would it make any difference in the gas bill if the meter were in the attic instead of in the cellar? In apartment houses with separate meters for each apartment, do the people on the top floor get more or less gas for their money?

CHAPTER V

NON-PARALLEL FORCES

Representation of forces by arrows — the parallelogram of forces — composition and resolution of forces — application to roof truss, friction, sailboat, and aëroplane.

106. Three forces acting at a point. In machines and other contrivances it often happens that forces which are not parallel balance each other and are thus in equilibrium. For example, suppose a street lamp is suspended over a street by a wire stretched between two posts, as shown in figure 110. We have here three non-parallel forces in equilibrium — first, the vertical pull OW due to the weight of the lamp; second, the pull exerted by one of the ropes; and third, the pull exerted by the other rope. We are to find what relation must exist between the magnitude and direction of any three such forces, if they are to produce equilibrium.

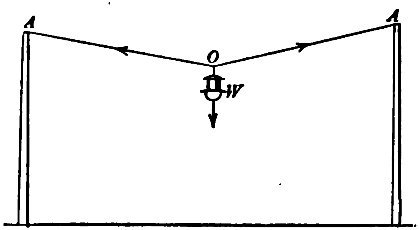


FIG. 110. — Three non-parallel forces.

107. Representation of forces by arrows. It will help us to form a mental picture of these three forces if we represent them by three arrows. The direction of each force will be indicated by the direction of the arrow, the point of application by the tail of the arrow, and the magnitude of the force by the length of the arrow, drawn to some convenient scale.

Thus in figure 111 we have an arrow 5 units long, and if we assume that each unit represents 10 pounds, the arrow *AB* shows a force of 50 pounds applied at *A*, acting due east. Figure 112 represents two forces, one *OA* of 30 pounds, acting due east applied at *O*, and the other *OB* of 40 pounds, acting due north, applied at the same point *O*.

FIG. 111.—A force of 50 pounds acting east.

If these two forces act simultaneously upon the body at *O*, the result will be the same as if a single force were applied, acting somewhere between *OA* and *OB*, but nearer the greater force *OB*. *This single force, which produces the same result as two forces, OB and OA, is called their resultant.*

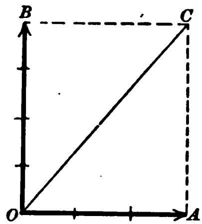


FIG. 112.—Resultant of two forces at right angles.

108. Principle of parallelogram of forces. If a **parallelogram** is constructed on *OA* and *OB*, the **diagonal** *OC* represents the **resultant**, as can be proved by the following experiment.

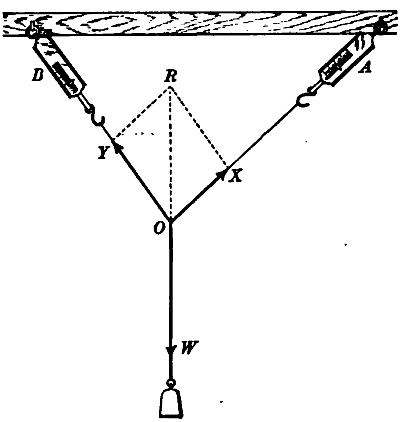


FIG. 113.—Experiment to illustrate parallelogram law.

Suppose we hang two spring balances *A* and *B* from two nails in the molding at the top of the blackboard, as shown in figure 113, and tie some known weight *W* near the middle of a string joining the hooks of the two balances. If we draw lines on the blackboard behind each of the three strings, we shall have represented the direction of each of the three forces. Then, if we note the tension in each string as shown by the amount of the weight *W* and the readings of the spring balances *A* and *B*, we may remove the apparatus and complete the diagram. Choosing some convenient scale, we

measure off on OA a distance corresponding to the tension in OA, and place an arrowhead at X, and in the same way we locate Y on OB. Then we construct a parallelogram on OX and OY by drawing XR parallel to OY and YR parallel to OX. It will be evident that the diagonal OR is the resultant of OX and OY, for if we measure OR and determine its magnitude from our scale of force, we find that this resultant OR is almost exactly equal and opposite to the third force OW. That is, either OR, or OX and OY, balances OW.

The force necessary to balance or hold in equilibrium two forces is called the **equilibrant.** Thus in the case just described, the force OW is the equilibrant of the two forces OX and OY.

The resultant of two forces acting at any angle may be represented by the diagonal of a parallelogram constructed on two arrows representing the two forces.

When three forces are in equilibrium, the resultant of any two of the forces is equal and opposite to the third, which can be regarded as their equilibrant.

109. Resultant depends on the angle between components. To determine the resultant of two or more forces, we must know not only the magnitude of the "components," but also the angle between them. This will be made clear by studying the same two forces at different angles, as in figure 114. It will be seen that the resultant OR gradually increases as

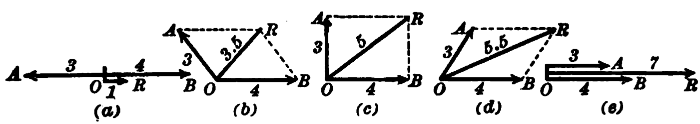


FIG. 114. — Two forces at varying angles.

the angle between the components OA and OB decreases. For example, if the angle is 180° (case (a)), the forces OA and OB are opposite and the resultant is the difference between the forces, $4 - 3$, or 1, and acts in the direction of the greater force, *i.e.* toward the right. As the angle gradually decreases the resultant OR increases, until, when the angle is

0° (case (*e*)), the forces *OA* and *OB* are acting in the same straight line and in the same direction, and the resultant is the sum of the two forces, 4 + 3, or 7. When the forces are at right angles (case (*c*)), the resultant can be computed from the geometrical proposition about the sides of a right triangle, namely, *the square on the hypothenuse is equal to the sum of the squares on the two sides.*

Thus,
$$\overline{OR}^2 = \overline{OA}^2 + \overline{OB}^2,$$
$$\overline{OR}^2 = 3^2 + 4^2 = 25,$$
$$OR = 5.$$

For oblique angles, such as (*b*) and (*d*) in figure 114, the resultant can be determined by plotting the forces to scale, or by trigonometry.

The process of finding the resultant of two or more component forces is called the **composition of forces.**

110. Composition of forces. — Illustrative Examples. Suppose we have a crane arm *AB*, attached to the wall as shown in figure 115 (*a*). The weight *W* is 2000 pounds and the tension in the cable *AC* is 1500 pounds. What is the force exerted by *AB*? In the solution of such problems it will be found helpful to draw a **force diagram** (Fig. 115 (*b*)), where *AW* represents the pull of the weight *W*, *AC* the pull of the cable, and *AE* the thrust of the crane arm. As these three forces are in equilibrium, we can apply the principle of the parallelogram of forces. We want to find *AR*, the resultant of *AC* and *AW*. Since *AC* and *AW* form a right angle, we know that

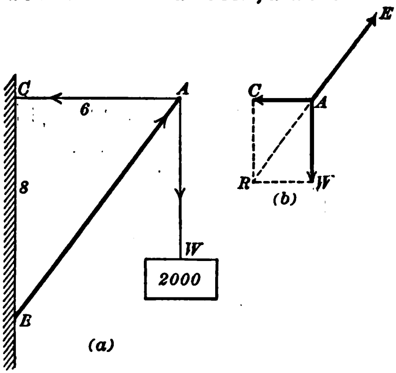


Fig. 115. — Three forces acting on crane.

$$\overline{AR}^2 = \overline{AC}^2 + \overline{AW}^2 = \overline{1500}^2 + \overline{2000}^2,$$
or
$$AE = 2500 \text{ pounds.}$$

Therefore the push exerted by *AB* is 2500 pounds.

Again, suppose we have a 100-pound child in a swing (Fig. 116). A man pushes the child to one side with a force of 20 pounds. What is the magnitude and direction of the pull exerted by the rope? In the force diagram (Fig. 116), CW represents the weight of the child (100 pounds), CP represents the push (20 pounds) of the man against the child, and CR represents the pull of the rope which we wish to determine. The resultant CR' of CP and CW is equal to $\sqrt{CP^2 + CW^2}$, or $\sqrt{(20)^2 + (100)^2}$, or about 102 pounds. Therefore the tension in the rope is also 102 pounds. Its direction can be found from the diagram.

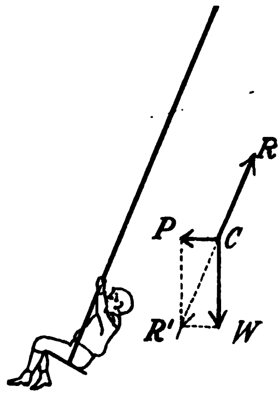


Fig. 116. — Three forces acting on child in swing.

Problems

1. Find, by plotting to scale, the resultant of a force of 8 pounds toward the east, and one of 4 pounds toward the north.
2. Compute the resultant in problem 1.
3. A force of 100 pounds acts north and an equal force acts west. What is the direction and magnitude of the equilibrant?
4. Find the resultant of a force of 10 pounds east and one of 14 pounds southwest.
5. Two forces, 5 pounds and 12 pounds, act at the same point. Find their equilibrant, (*a*) if they act in the same direction, (*b*) if they act in opposite directions, and (*c*) if they act at right angles.

111. Resolution of forces. The principle of the composition of forces can be worked backward. If one force is given, we can find two others in given directions which will balance it. For example, take the case of the lamp suspended above the middle of the street. If we know the weight of the lamp and the angle of sag of the ropes, as shown in figure 117, we can calculate the tension in the ropes.

Suppose that the weight of the lamp is 50 pounds, and that the rope ALB sags so as to make both the angle ALR and the angle BLR equal

to 75°. In the diagram (Fig. 117) draw the arrow LW down from L to represent 50 pounds on some convenient scale. As the two ropes have to hold up the lamp, the resultant of the forces representing the tension in the ropes must be equal and opposite to the force representing the weight. So we draw LR equal and opposite to LW. Then we construct a parallelogram on LR as a diagonal with its sides parallel to LA and LB, Ry being drawn parallel to LA, and Rx parallel to LB. Ly represents the tension in the rope LB and is equal to about 96.6 pounds, and Lx represents the tension in LA and is also equal to about 96.6 pounds.

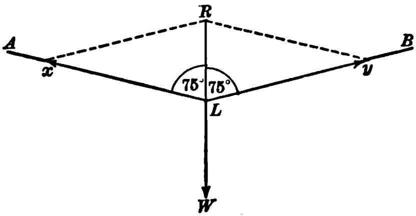


FIG. 117.—Three forces acting on street lamp.

Another good example of the resolution of one force into two forces which just balance it, is the case of a street lamp hung out on a bracket from a pole, as shown in figure 118 (*a*).

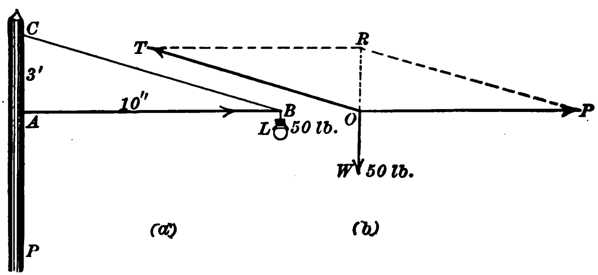


FIG. 118.—Three forces acting on lamp hung on bracket.

Suppose the lamp L, weighing 50 pounds, is hung out from a pole PC by means of a stiff rod AB, 10 feet long, and a tie rope or wire BC, which is fastened to the pole at C, 3 feet above A. What is the force exerted by the rope BC?

In the diagram (Fig. 118 (*b*)), the weight of the lamp is represented by OW, the push of the rod AB by OP, and the tension of the tie rope BC by OT. Since we know the force OW (50 pounds), we draw this line to some convenient scale. The resultant of OP and OT must be equal and opposite to OW. Therefore we make OR equal and opposite to OW.

Then, completing a parallelogram on OR as a diagonal, we have OP representing the push of the rod against the lamp, and OT the tension in the tie rope BC. If we draw these lines carefully to scale, we find that the tension is 174 pounds.

In general, a single force may be resolved into two components acting in given directions, by constructing a parallelogram whose diagonal represents the given force, and whose sides have the given directions of the components.

112. Component of a force in a given direction. If a force is given, we can find two other forces, one of which represents the **whole effect in a given direction** of the given force.

Thus in figure 119 we have a canal boat AB which is being towed by the rope BC. We may resolve the force along the rope BC into two components, one of which, BE, is

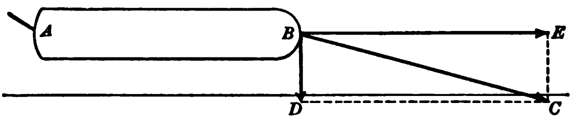


FIG. 119. — Useful component of force on canal boat.

effective in pulling the boat along the canal, and the other, BD, at right angles, is useless or worse than useless, since it tends to pull the boat toward the bank. BE, the useful component of BC, can be computed by drawing the force BC to scale and then constructing a **rectangle** on BC as a diagonal, such as $BECD$.

113. Applications of the principle of parallelogram of forces. This principle is one of the foundation stones in the study of mechanics. When stated with the aid of a geometrical diagram, it seems simple, but when met in a crane, derrick, bridge, or roof truss, it is puzzling. This is because, in solving practical problems, we seldom find bodies which are small enough to be regarded as points at which forces act. Nevertheless we can do problems by this method, even when the bodies are quite large. For if any body is held still by

three forces, their lines of action, if prolonged, must go through a single point, as shown in figure 120 (*a*). If this were not true of three forces acting on a body (Fig. 120 (*b*)), it would *spin* around. So we can think of the forces as acting at a single point, even though the body in which the point lies is quite large.

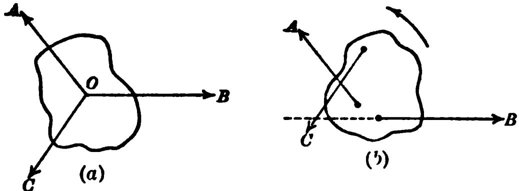


FIG. 120. — Condition of spin and rest.

114. Roof truss. When a wooden house is built, the roof is usually supported by pairs of timbers set like an inverted V, as in figure 121. Each pair of timbers has to carry the weight of a section of the roof, and, in winter, of the snow and ice that accumulate on it. This weight is really distributed along the timbers, but it can be thought of as concentrated, half at the peak and half at the eaves, where it rests directly on the walls. The part of the load that is at the peak tends to "spread" the inverted V, and our problem is to find what has to be done to prevent this.

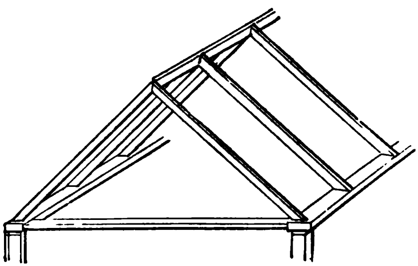


FIG. 121. — Roof trusses.

We may test this experimentally with a small model of a pair of roof trusses (Fig. 122). These have hinges at the top instead of a stiff joint, and frictionless wheels underneath, so that they will not stand up at all under the load *W* unless a tie is put across the bottom of the Λ to prevent the spreading. If a spring balance is put into the tie, the pull which the tie has to exert on the truss members can be measured. If the load at the peak is 50 pounds, and if the truss members make a right angle, the "tension" in the tie will be about 25 pounds.

In discussing this experiment, we have to apply the parallelogram of forces at two points separately. In the first place, let us consider the pin of the hinge at the top. This is acted on by three forces, the pull of the weight W, and the push exerted by each rod (Fig. 122). Since these balance, we can find each push by constructing a parallelogram whose diagonal is equal and opposite to W. If the rods are at right angles, this parallelogram is a square, and each push is $50/\sqrt{2}$, or 35.3 pounds.

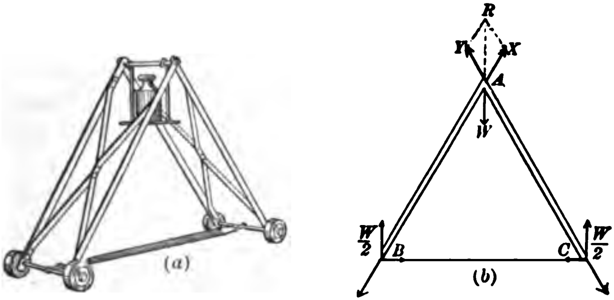


FIG. 122.— Experimental roof truss, and force diagram.

Turning next to the pin at the foot of one rod, we see that it is also acted on by three forces, the push of the rod, 35.3 pounds, the pull of the tie wire, and the push upward exerted by the table. Since these balance, we can get the last two by constructing a parallelogram on the known force as a diagonal (draw this yourself). This parallelogram is composed of two 45° triangles, and so both the pull of the tie wire and the push of the table are $35.3/\sqrt{2}$, or 25 pounds.

In building a roof, the pull exerted by the tie wire in our experiment has to be provided for in some way. Usually the ends of the roof timbers are nailed to the frame of the building, which is stiff enough to exert a part or all of the required force. Often a board is nailed across the inverted V, either at the bottom, or a little higher up, to help exert it. In large roof trusses, as in churches, an iron rod is strung across and tightened with a screw coupling.

115. Bridge. Large bridges are built of wood or steel "members" joined to form a number of adjacent triangles. If the members are strong enough not to stretch or shrink

under the loads imposed on them, each triangle, having three sides of unchanging length, keeps its shape, and so the whole truss is rigid.

In very large bridges the members are joined together at the corners of the triangles by boring holes in them and thrusting a steel pin through all the holes at a joint. Bridges made in this way are called **pinned bridges.** The plate opposite page 114 shows two such bridges. In designing a pinned bridge, an engineer computes the "stresses in the members," that is, the forces which they have to exert to hold the bridge stiff under load, by applying the parallelogram of forces to the pin at each joint separately. The members which have to push against the pins at their ends are called **compression members,** because they tend to shorten under load, while those that have to pull on the pins at their ends are **tension members,** and tend to lengthen under load. In large bridges it is easy to see which are compression members and which tension, for the compression members are made broad and stiff with "latticing" up their sides, while the tension members are steel straps or rods with enlarged ends to give room for the holes. Thus the heavy lines in figure 123 indicate compression members, while the light lines correspond to tension members.

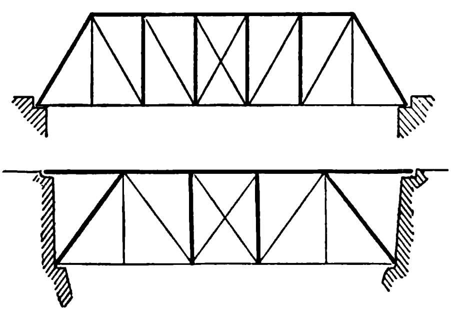


FIG. 123.— Diagrams of the bridges in the plate opposite page 114.

In smaller bridges the members are not joined by pins, but are **riveted** to "gusset plates" at each joint. Such bridges are designed as if they were pinned, the stiff joints giving an additional factor of safety.

Framed bridges with pinned joints. The upper one is a "through-bridge" with 9 "panels," the lower a "deck-bridge" with only 5 "panels." Notice, however, that the essential features of the two trusses are alike.

The smallest steel bridges are supported by **plate girders**, one on each side, which are simply stiff steel beams, and will be discussed in the next chapter.

Roofs of large span are often supported by framed trusses, made of members forming triangles, like bridge trusses.

116. How a boat sails into the wind. Let AB (Fig. 124) represent a boat, SS' its sail, and W the direction of the wind. It is sometimes hard to see how such a wind pushes the boat ahead instead of forcing it backward. The wind blowing it against a slanting sail SS' is deflected and causes a pressure perpendicular to the surface. This pressure can be represented by the arrow cP in the diagram. The force cP can be resolved into two components, one useful, ck, which is parallel to the keel of the boat, and the other useless, cw, which tends to move the boat to leeward. This sideways movement is largely prevented by a deep keel or a centerboard. So the net effect of the wind is to drive the boat forward.

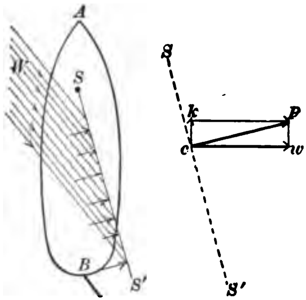


FIG. 124. — Action of wind on a sail.

117. What supports an aëroplane? An aëroplane of the monoplane type has one huge plane or sail, and a biplane two such planes, one above the other. These planes are tilted so that the front edge is a little higher than the rear edge. There is also a light but powerful gasolene engine which, by turning one or two large propellers, forces the aëroplane forward. How such a machine, which is heavier than air, is kept up, will be seen from figure 125. Let AB represent a tilted plane moving from right to left. The conditions are evidently the same as if the plane stood still and a strong wind was blowing, as shown by the arrows. The air striking against the under side of the plane AB is deflected and

causes a pressure OP at right angles to the surface. It is the upward component of this force which keeps the aëroplane from falling. The weight of the machine, including the engine and load, is represented by OW. The driving force of the propeller OT must be equal to the resultant of the two forces OP and OW to keep the machine from slowing up.

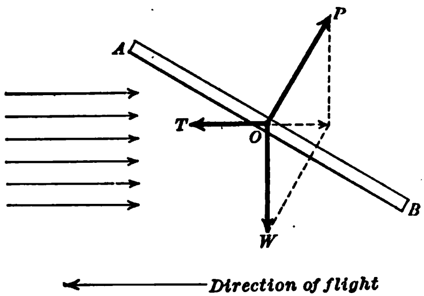


FIG. 125.—Forces acting on aëroplane.

118. Friction on an inclined plane. When an object is placed on an inclined plane, friction tends to keep the object from sliding down the plane (Fig. 126). If the angle of inclination is small enough, this friction will prevent the object from sliding down the plane.

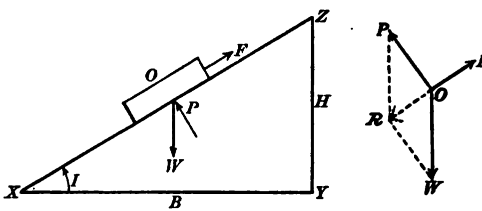


FIG. 126.—Friction on inclined plane.

For example, suppose an electric car is on a grade with the brakes set, so that the car stands still. How steep can the grade be before the car slides down? In the diagram, figure 126, let OW represent the weight of the car, OP the pressure of the inclined plane against the car, and OF the friction which retards its motion. When these three forces are in equilibrium, the resultant of OP and OW, that is, OR, must be opposed by an equal force OF. Now OF can never exceed a limiting value which depends on the pressure and on the coefficient of friction, the latter being determined by the condition of the track. But the resultant OR increases as

the incline becomes steeper. So, as the steepness increases, we soon reach a condition in which OR is greater than OF possibly can be, and the car slides down. If we know the coefficient of friction between the wheels and the rails, we can compute the grade at which the car will begin to slide.

Let figure 126 represent this grade. We have already (section 45) defined the coefficient of friction as the ratio between friction and pressure, and so, in this case, we have

$$\text{Coefficient of friction} = \frac{OF}{OP} = \frac{OR}{OP}.$$

From geometry we know that the triangles OPR and XYZ are similar since they are mutually equiangular. It follows that

$$\frac{OR}{OP} = \frac{H}{B} = \frac{\text{height of plane}}{\text{base of plane}}.$$

Therefore

$$\textbf{Coefficient of friction} = \frac{\textbf{height of plane}}{\textbf{base of plane}}.$$

This is a convenient way of measuring coefficients of friction.

Problems

1. If the resultant of two components acting at right angles is 50 pounds, and one of the forces is 15 pounds, what is the other force?

2. One of two components acting at right angles is three times the other. Their resultant is 32 pounds. Find the forces.

3. A force of 8 pounds is to be resolved into two forces, one of which is 12 pounds, and makes an angle of 90° with the given force. Find the other force.

4. A boy weighing 50 kilograms sits in a hammock whose ropes make angles of 30° and 60°, respectively, with the vertical. What is the tension in each rope?

5. Each rope in problem 4 is fastened to a hook in the ceiling. Find the vertical pull on each hook.

6. Figure 127 shows a simple crane. Find the tension in the tie rope BC and the push of the brace AC, when the weight W is one ton, and the angle BAC is 45°.

7. In figure 119 the canal boat is 10 feet from the shore, and a pull of 200 pounds is exerted on the 50-foot towline. What is the effective component?

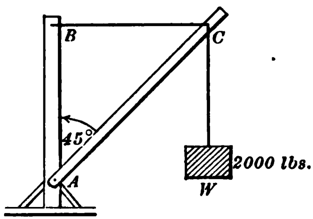


FIG. 127. — Diagram of simple crane.

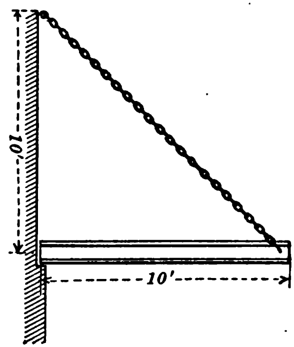


FIG. 128. — Girder supported from a wall.

8. One end of a horizontal steel girder 10 feet long rests on a ledge in the wall, and the other end is supported by a steel cable arranged as shown in figure 128. Assuming that the girder weighs 40 pounds per foot, find the tension in the cable.

SUMMARY OF PRINCIPLES IN CHAPTER V

Forces can be represented by arrows.

The parallelogram of forces:—

The *resultant* of two forces is the *diagonal* of their parallelogram.
The *equilibrant* of two forces is equal and opposite to their resultant.

If three forces act on a body (not a point), their lines of action must pass through a single point, and the parallelogram principle can be used.

QUESTIONS

1. Show by a diagram the useful component of the pull exerted on a sled by a rope.

2. Why is a long towline more effective in hauling a canal boat than a short line?

3. Why does one lower the handle in pushing a lawn mower through tall grass?

4. A boat is rowed across a river. What two forces are acting on the boat?

5. A child sitting in a swing is drawn gradually aside by a force which continually acts in a horizontal direction. Does the tension in the swing rope grow smaller or larger?

6. Why will a long rope hanging between two points at the same level break before it can be pulled tight enough to be straight?

7. Find in some building a roof truss with a steel tie rod to keep it from spreading.

8. How are the walls of Gothic cathedrals strengthened so that they can exert the side thrust necessary to hold up the roof?

9. Examine the steel bridges in your neighborhood to see if they are "girder bridges" or "framed bridges," and, if any of them are framed, see whether they are pinned or riveted, and which members are compression members, and which tension. Make a sketch like those in figure 123 of one of these bridges, showing the compression members by heavy lines.

10. Explain by diagram how an ice boat may go faster than the wind.

11. Could an ordinary balloon "tack" against the wind like a sailboat, if it was provided with a sail, a large keel, and a rudder, like a sailboat? Why?

CHAPTER VI

ELASTICITY AND STRENGTH OF MATERIALS

The different kinds of stress — stress and strain — Hooke's law — elastic limit — breaking strength — factor of safety.

Unit stress and unit strain in tension — tensile strength — stiffness and strength of beams — cross sections of beams.

119. Importance of studying materials. A structural engineer who is to build a bridge, a building, or a machine must know not only the forces that will be exerted on each of its parts, but also the strength of the wood, brick, stone, or steel of which they are to be made. This knowledge can be gained only by testing each kind of material with the greatest care. For this reason, every engineering handbook tabulates the results of a great number of tests of this kind. Every large manufacturer of steel girders or rails maintains a testing laboratory so that he can sell his products under a strength guarantee. Even textile manufacturers test the breaking strength of the yarn that goes into their cloth. Indeed, the study of the properties of structural materials is regarded as of such importance to the public that the government itself maintains a bureau for the purpose. In this chapter we shall learn how to make such tests on a small scale, and how the results are used.

120. The different kinds of stresses. In designing a beam or column, or some part of a machine, an engineer must first know how the force it is to resist will be applied.

For instance, the cable that supports an elevator, or the rope of a swing, or a belt that is transmitting power from one pulley to another, has to resist a pull applied at each

end, which tends to stretch it, and may, perhaps, break it by pulling one part of it away from the next. In such a case we say that the "member" — that is, the cable, or rope, or belt — is in **tension**, meaning "in a state of tension."

The pier of a bridge, or the foundation of a house, or a post supporting a piazza roof, has to do something quite different from this. It has to resist a push at each end, which tends to shorten it, and may cause it to give way by crushing it. In such a case we say that the member — that is, the pier, or foundation, or post — is in **compression**, meaning "in a state of compression."

A floor beam in a house or a girder in a plate-girder bridge (section 115) has to resist **bending**, and if it gives way at all, it does so by breaking in two like a stick broken across one's knee.

The duty of the shaft that drives the propeller of a steamship, or of the shafts that run overhead in many factories and transmit power to the various machines, is to resist **twisting**.

And, finally, the duty of a rivet in a steel structure is different from any of these (see Fig. 129). It has to keep one of the plates from sliding over the other. When such a rivet gives way, it is often because the halves of it have been pushed sidewise so hard that one has slid away from the other, leaving a flat, clean break parallel to the surface separating the plates. It is a strain of this sort that we are really putting on a piece of cloth or paper when we cut it with a pair of shears. So we say that the rivet is in **shear**, meaning that it is in the same state as if it were being cut in two by a pair of huge shears.

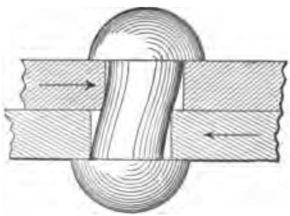


FIG. 129. — Action of plates on rivet.

There are, then, these five kinds of stresses: tension,

compression, bending, twisting, and shear. In each case that material should be used which will best resist the particular kind of stress that is to be applied to it. Thus bricks set in mortar do very well under compression, but are of little use in resisting any of the other kinds of stress. Steel will resist any of them well. Cast iron will resist compression about four times as well as it will tension, and so on.

121. Stress and strain. Whenever any one of these kinds of stress is applied to a body, the body yields a little. No bridge girder is stiff enough not to bend a little under every wagon or train that goes over the bridge. If it is a good girder, the amount of bending is imperceptible to ordinary observation; but there is always some bending. Similarly, every shaft on a steamship twists a little when the propeller is in motion. Measuring this very small twist is often the only way in which the horse power delivered by the engine to the propeller can be measured. The same can be said of the other types of stress; each of them always causes some yielding or deformation of the body under stress.

The word "strain" is used to describe the deformation produced. The word **stress** always refers to the forces which are acting, while the word **strain** refers to the effect which they produce.

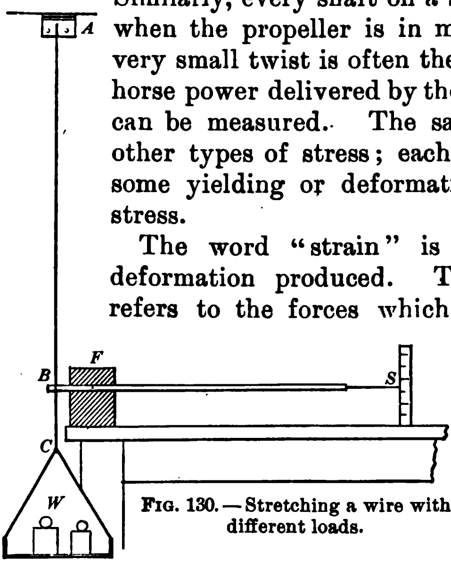


Fig. 130. — Stretching a wire with different loads.

122. Relation of strain to stress. Let us try some experiments to see if there is any relation between the amount of stress applied to a body and the amount of strain it produces.

I. *Tension.* Let us fasten one end of a piece of steel or spring-brass wire in a clamp near the ceiling, and attach a pan for weights to the lower end of the wire (Fig. 130). Since the stretch will be small, it is necessary to use a lever or some other device to magnify it. Having placed just enough weight in the pan to straighten the wire, we add weights one at a time and read the corresponding positions of the pointer. Each time we must remove the added weights to see if the pointer comes back to its original position. When it fails to do this, we will stop the experiment and disregard, for the moment, the last reading of the pointer. If we then compute from each deflection of the pointer the actual stretch or elongation of the wire, and divide each stretch by the force causing it, we will find that all the quotients are approximately the same. That is, the stretch is proportional to the load.

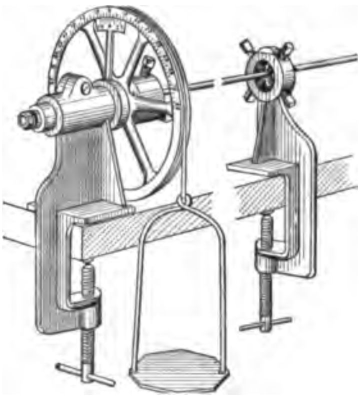


FIG. 131. — Twisting metal rods.

II. *Compression.* The same is true for compression. Thus experiments have shown that under ordinary conditions the compression of a spring is proportional to the force applied.

III. *Bending.* We can perform a similar experiment for bending by supporting a metal rod or tube on knife edges, and hanging different weights from the center. A lever, like that used in the tension experiment above, enables us to measure the small deflections of the center of the rod. As before, we find that the deflections are proportional to the loads causing them.

IV. *Twisting.* The apparatus shown in figure 131 enables us to perform similar experiments on twisting. As before, we find that the twist is proportional to the stress causing it, namely, the "torque" or moment of the twisting force.

In all these cases, *the strain is proportional to the stress.* This is called **Hooke's law**, after the medieval scientist who discovered it. Hooke's law applies to all kinds of strains, if the stresses are not too great.

Problems

1. If a weight of 1 pound, when hung on a certain spring, lengthens it 2 inches, what weight would lengthen it $\frac{1}{2}$ an inch? How much would $\frac{3}{8}$ of a pound lengthen it?

2. If a force of 5 pounds is required to move the middle point of a beam $\frac{1}{5}$ of an inch, what force would move it $\frac{1}{2}$ an inch?

3. A 2 pound force is applied to the rim of a wheel 9 inches in diameter in the torsion apparatus described in section 122, and the end of the rod twists through 3°. What force would have to be applied to the rim of a wheel 12 inches in diameter to make the end of the same rod twist through 5°?

123. Elastic limit and breaking strength. In the tension experiment in the last section we found that when a sufficiently great load was hung from the wire, the latter did not shrink back to its original length when the load was removed. It had acquired a permanent "set." The same thing is true of other kinds of stress, and might have been noticed in the other experiments. The smallest stress of any particular kind that will cause a permanent set in a body is called the **elastic limit** of the body for that particular kind of stress. As long as the load is below the elastic limit, Hooke's law holds, but stresses greater than the elastic limit cause deflections greater than Hooke's law predicts.

If we still further increase the load in the tension experiment, we finally reach a load so great that the wire stretches very rapidly and almost immediately breaks. This is also true of other kinds of tests, such as tests for bending. The smallest stress of any particular kind that will cause a body to give way is called the ultimate or **breaking strength** of the body for that particular kind of stress.

Usually the elastic limit of anything is much smaller than its breaking strength. But certain materials, such as glass, follow Hooke's law right up to their breaking points, and never show a permanent set. In such cases, the elastic limit and the breaking strength are equal.

124. Factor of safety. An engineer, when designing a bridge or a machine, must be absolutely sure that no part of it will ever be subjected to a stress greater than its elastic limit, for if this were to happen, the part would be permanently deformed, and this would weaken the rest of the structure, or at least throw it out of alignment. He therefore plans to make each member big enough to carry several times as much load as will probably ever be imposed on it. This is partly to provide for any unforeseen temporary overloading of the structure, and partly because there may be, even in materials of the best quality, imperceptible flaws that would make the completed member less strong than it seems to be. The number of times that the load planned for is greater than the load expected is called the **factor of safety**.

The factor that should be used varies with the material; thus it is commonly 10 for brick and stone and only 4 for steel. It also varies with the nature of the load; thus it is commonly larger when the load is to be intermittent, as in machines or railroad bridges, than when it is to be steady, as in buildings. Often the factor for buildings is taken larger than would otherwise be necessary, so that there may be no danger of deflections in the walls and ceilings great enough to crack the plaster.

125. Unit stress and unit strain in tension. When we discussed Hooke's law in section 122, we were comparing with each other the deformations produced in *the same* wire or rod by forces of different magnitudes. That is, if we knew by experiment how much one force would stretch a given wire, we could compute how much a different force would stretch *the same* wire. Let us now see if we can compute from the result of an experiment on *one* wire how much *another* wire of the same material but of a different shape would be stretched by any force that might be applied to it. This is important for the engineer because it enables him to

test a small piece of a particular kind of steel, and compute from this how a large tension bar in a bridge will act. For this purpose it will be convenient to define more precisely than in section 121 the meaning of "stress" and "strain" in the case of a wire or rod under tension.

If one wire is twice as long as another, a given pull will stretch the long wire twice as much as the short one; for each half of the long wire is just like the whole of the short one, and has to pull just as hard on its supports; each half, then, stretches as much as the whole of the short wire. In general, the total stretch of a wire under a given load is proportional to its length. The *stretch per unit of length* is called the **unit stretch** or **unit strain.**

$$\textbf{Unit stretch} = \frac{\textbf{total stretch}}{\textbf{length}}.$$

For example, a piece of steel piano wire, originally 90 inches long, is stretched 0.033 inches by a certain load. Then the unit stretch or unit strain is 0.033/90 or 0.00037 inches per inch of length.

Similarly if two wires are of the same length, but one has twice as great an area of cross section as the other, the thick wire is equivalent to two of the thin wires side by side, and it would take twice as much force to stretch the thick wire a given amount as to stretch the thin wire the same amount. In general, the pull required to produce a given stretch will be proportional to the area of cross section. The *pull per unit area of cross section* is called the **unit pull** or **unit stress.**

$$\textbf{Unit pull} = \frac{\textbf{total pull}}{\textbf{area of cross section}}.$$

For example, a piece of steel piano wire 0.0348 inches in diameter is subjected to a pull of 10 pounds. What is the unit pull or unit stress in the wire? The area of the cross section of the wire is πr^2 or 3.14×0.0174^2 or 0.000950 square inches. Therefore the unit stress is 10/0.000950 or 10,500 pounds per square inch.

If we were to test with the apparatus described in section 122 a number of wires of the same material but of different

sizes and lengths, we would find that in all cases *the unit stretch is proportional to the unit pull.* This law enables us to compute how much a force will stretch a wire, if we know how much another force will stretch another wire of the same material.

For example, if a 10-kilogram weight produces in a piece of piano wire, 0.5 millimeters in diameter and 1 meter long, a stretch of 0.02 millimeters, what will be the stretch produced in a piece of piano wire 0.4 millimeters in diameter and 2 meters long, by a 15-kilogram weight?

For the first wire: —

The cross section is $\pi r^2 = \frac{\pi}{16}$ square millimeters.

The unit pull is $10 \div \frac{\pi}{16} = \frac{160}{\pi}$ kilograms per square millimeter.

The unit stretch is 0.02 millimeters per meter.

For the second wire: —

The cross section is $\pi r^2 = \frac{\pi}{25}$ square millimeters.

The unit pull is $15 \div \frac{\pi}{25} = \frac{375}{\pi}$ kilograms per square millimeter.

Call the unit stretch x millimeters per meter.

Then, since the unit pulls and unit stretches are in proportion,

$$\frac{x}{0.02} = \frac{\frac{375}{\pi}}{\frac{160}{\pi}},$$

or

$$x = 0.02 \times \frac{375}{160} = 0.047 \text{ millimeters per meter.}$$

Since the second wire is 2 meters long, the total stretch in it is

$$2 \times 0.047 = 0.094 \text{ millimeters.}$$

126. Tensile strength. The length of a wire or rod has nothing to do with its strength under tension, unless the rod is so long that its own weight has to be taken into account. The strength of a wire or rod is proportional to the area of its cross section. The strength of a wire or rod of unit cross section (1 square inch or 1 square centimeter) is called the

tensile strength *of the material.* Tables giving the breaking strengths of various materials can be found in any engineer's handbook.

For example, in a testing laboratory it was found that a wrought-iron bar 0.75 inches in diameter broke under a pull of 28,700 pounds. The tensile strength of the material was, then,

$$\frac{28700}{3.14 \times (0.375)^2} = 65{,}000 \text{ pounds per square inch.}$$

PROBLEMS

1. If a pull of 22 pounds will break iron wire of size 24, what pull will break iron wire of size 30? (See table on page 304 for diameters.)
2. From the data given in problem 1, compute the tensile strength of the iron.
3. An iron bar is to be subjected to a total pull of 35,000 pounds and is to be designed so that the unit pull shall not exceed 2500 pounds per square inch. What should be the area of its cross section, and if round, what should be its diameter?
4. A force of 2 kilograms stretches a certain wire 3 millimeters. How much will a force of 5 kilograms stretch the same wire?
5. How much would 5 kilograms stretch a piece of the same kind of wire as in problem 4, with the same diameter but twice as long?
6. How much would 5 kilograms stretch a piece of the same kind of wire as in problem 4, of the same length but with twice the diameter?
7. How much would 5 kilograms stretch a piece of the same kind of wire as in problem 4, but with half the diameter and three times the length?

127. Stiffness and strength of beams. The design of floor beams for buildings or girders for bridges is another matter in which it is important for engineers to be able to predict from experiments on small test pieces how a full-sized member will act. They have, therefore, tried many experiments on beams of different sizes and shapes with large test machines similar in principle to the bending apparatus described in section 122. These experiments have shown that what may be called the **stiffness factor** of a beam of rectangular cross section is

$$\textbf{Stiffness factor} = \frac{\textbf{breadth} \times (\textbf{depth})^3}{(\textbf{length})^3}.$$

To compare the deflections that would be produced by a given force in two beams, we have only to compute their stiffness factors, and the one with the larger stiffness factor will bend less under the given load.

For example, which of the following beams is stiffer:—
Beam A: length 10 feet, breadth 4 inches, depth 6 inches;
Beam B: length 20 feet, breadth 6 inches, depth 8 inches?

$$\frac{\text{Stiffness factor of } A}{\text{Stiffness factor of } B} = \frac{4 \times 6^3 \div 10^3}{6 \times 8^3 \div 20^3} = \frac{9}{4} \text{ (by cancellation).}$$

Therefore, beam A is more than twice as stiff as beam B; that is, if equal weights were hung from the centers of the two beams, the center of A would drop less than half as far as would the center of B.

Experiments have shown, however, that the stiffer of two beams is not necessarily the stronger. In fact, the **strength factor** of a beam of rectangular cross section is quite different from its stiffness factor. It is

$$\text{Strength factor} = \frac{\text{breadth} \times (\text{depth})^2}{\text{length}}.$$

For example, compare the strengths of the two beams described above.

$$\frac{\text{Strength factor of } A}{\text{Strength factor of } B} = \frac{4 \times 6^2 \div 10}{6 \times 8^2 \div 20} = \frac{3}{4}.$$

Therefore, although beam A is more than twice as stiff as beam B, it could support only three quarters as much load without breaking.

128. Cross sections of beams. Wooden beams ordinarily have a rectangular cross section, and are designed on the basis of the laws in the last section; but steel beams, if so designed, would be too heavy. It is possible, however, to distribute a much smaller amount of material in such a way as to be just as effective. Thus we all know that a bicycle frame made of thin tubing is much stiffer than a frame *of equal weight* made of solid rods. So it pays to consider what the different parts of the cross section of a beam have to do to resist bending.

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$$\textbf{Stiffness factor} = \frac{\textbf{breadth} \times (\textbf{depth})^3}{(\textbf{length})^3}.$$

To compare the deflections [illegible] given force in two beams, we [illegible] stiffness factors, and the one with the larger [illegible] will bend less under the given load.

For example, which of the following [illegible]
Beam A: length [illegible]
Beam B: length [illegible]

[illegible]

Therefore, beam A is more than twice as stiff [illegible] equal weights were hung from the centers of [illegible] of A would drop less than half as far as would [illegible]

Experiments have shown, however, [illegible] beams is not necessarily the stronger. [illegible] factor of a beam of rectangular cross section [illegible] from its stiffness factor. It is

$$\text{Strength factor} = \frac{[illegible]}{[illegible]}$$

For example, compare the strengths of [illegible]

[illegible]

Therefore, although beam A is more [illegible] could support only three quarters as much [illegible]

128. Cross sections of beams. Wooden beams ordinarily have a rectangular cross section, and are designed on the basis of the laws in the last section; but steel beams, if so designed, would be too heavy. It is possible, however, to distribute a much smaller amount of material in such a way as to be just as effective. Thus we all know that a bicycle frame made of thin tubing is much stiffer than a frame of *equal weight* made of solid rods. So it pays to consider what the different parts of the cross section of a beam [illegible] to do to [illegible]

6. Where is the elastic medium in the human body which prevents injury to the brain when we jump?

7. In the frame of a bicycle, why does a pound of steel give greater stiffness in the form of tubing than in rods?

8. Which flange of a cast-iron girder should have a greater cross section? Notice the statement in section 120.

9. Try to find out what is meant by the "fatigue" of metals (see encyclopedia).

10. What advantages has reënforced concrete over ordinary concrete for building purposes?

11. How are the walls of high office buildings supported, and why?

CHAPTER VII

ACCELERATED MOTION

Speed and acceleration — laws of motion at constant acceleration — falling is motion at constant acceleration — value of acceleration of gravity.

129. Average speed. If a man walks 12 miles in 3 hours, we say that he averages 4 miles an hour. To be sure, at any particular point on his journey he may have been going faster or slower, but his average speed or velocity is 4 miles an hour. If we know that the average speed of a steamer is 22 miles an hour, we can find a day's run by multiplying the average speed by the number of hours in a day; thus $22 \times 24 = 528$ miles. In general,

Distance = average speed × time.

Speed is expressed in various ways; for example, we say that an automobile travels at the rate of 25 miles an hour, a steamer does 18 knots or 18 nautical miles an hour, a sprinter runs 100 yards in 10 seconds, and a rifle ball goes 2000 feet per second. For purposes of comparison it is convenient to have some uniform way of expressing speed, and so engineers and other scientific men have come to use feet per second (ft./sec.) or centimeters (or meters) per second (cm./sec. or m./sec.). The following table gives some average speeds: —

TABLE OF SPEEDS

Soldiers marching	4.3 ft./sec. =	1.4 m./sec.
Horse galloping	16 ft./sec. =	5.2 m./sec.
Ocean steamer	40 ft./sec. =	12.2 m./sec.
Express train	82 ft./sec. =	26.9 m./sec.
Wind in hurricane	165 ft./sec. =	54.2 m./sec.
Sound	1120 ft./sec. =	386 m./sec.
Rifle ball	1500 to 2000 ft./sec. =	493 to 657 m./sec.

Questions and Problems

1. Sixty miles an hour equals how many feet per second? You would do well to remember this number.

2. With the help of a time-table, compute the average speed of an express train, and of a local.

3. If the distance across the Atlantic Ocean is about 3000 miles, how many days will it take a steamer to cross, at the speed given in the table above?

4. An officer on horseback starts on the gallop to overtake his regiment a mile away, which is marching ahead. If they travel at the speeds given in the table, how long will it take him?

5. How long will it take an express train to cover 50 miles, going at the rate given in the table?

6. A rifle is fired at a target half a mile away. How long after it is fired does the sound it makes against the target reach the man with the rifle?

7. There is a common rule that if any one in a train counts for 19 seconds the number of clicks as the car passes over the ends of the rails, the number he gets will be the speed of the train in miles per hour. What must be the length of the rails to make this rule work?

130. Variable speed. When a train is starting out from a station, it is gaining speed, and when it is approaching a station where it must stop, it is losing speed. So we see that on account of stops and differences in grade, the speed of a train is not uniform or constant, but is changing or variable. When a loaded sled starts at the top of a long hill, it gains in speed as it descends the hill; but when it reaches the bottom, it is retarded and loses speed until it stops. Its speed or velocity, starting at zero, has increased to a maximum and then has decreased to zero again. Similarly, the speed of a projectile from a big gun or of the piston of an engine is not uniform but variable.

If we wished to determine the speed of an automobile at any instant or point, we would measure off some convenient distance near the point and then get the time which elapsed while the automobile traveled the fixed distance. For example, if the measured distance, sometimes called a "trap,"

was a quarter of a mile and the time was 20 seconds, the speed was three quarters of a mile per minute or 45 miles per hour. But if the driver of the automobile was aware of the trap and was driving at a dangerously high speed at the beginning of the trap, he would slow down so that his average speed over the measured distance would be within the limit. To catch such a driver, that is, to get his speed more accurately *at any point*, we take as short a distance as is consistent with an accurate measurement of the time.

131. Acceleration. It is unpleasant to be on a street car when it starts or stops too suddenly. This suggests the problem of measuring a **rate of change of speed**, which is called **acceleration.** It has been found that a city street car standing at rest can safely gain speed, so that at the end of 10 seconds it is going 15 miles per hour. Assuming that this gain in speed is made at a constant rate (only constant accelerations will be discussed in this book), the speed of the car increased 1.5 miles-per-hour every second. In other words, the acceleration was 1.5 miles-per-hour per second. Or, since 15 miles an hour is 22 feet per second, we can say that the gain in speed each second is 2.2 feet per second.

In general,

Acceleration = gain in speed per ***unit time,***

and acceleration is always to be expressed as so many **speed units per time unit.** Since there are many different speed units, such as miles-per-hour, kilometers-per-hour, feet-per-second, and centimeters-per-second, there are many ways of expressing the same acceleration. Thus the acceleration of the electric car just mentioned is

	Velocity Unit	Time Unit
	1.5 miles-per-hour	per second,
or	2.4 kilometers-per-hour	per second,
or	2.2 feet-per-second	per second,
or	67.0 centimeters-per-second	per second.

All these statements mean exactly the same thing. Engineers use the first two expressions for acceleration, while other scientific men more commonly use the last two. It is convenient to abbreviate "feet-per-second per second" as ft./sec.^{2} and "centimeters-per-second per second" as cm./sec.^{2}, but each of these abbreviated expressions means simply so many velocity units gained per second.

The accelerating rates of cars vary according to service and equipment, but the following rates are common in practical operation: —

Steam locomotive, freight service,	0.1–0.2 miles-per-hr. per sec.
Steam locomotive, passenger service,	0.2–0.5 miles-per-hr. per sec.
Electric locomotive, passenger service,	0.3–0.6 miles-per-hr. per sec.
Electric car, interurban service,	0.8–1.3 miles-per-hr. per sec.
Electric car, city service,	1.5 miles-per-hr. per sec.
Electric car, rapid transit service,	1.5–2.0 miles-per-hr. per sec.

132. Positive and negative acceleration. When the speed is **increasing**, the acceleration is said to be **positive**, and when the speed is **decreasing**, the acceleration is **negative.** Thus when a baseball is dropped from a tower, it goes faster and faster; it has *positive acceleration.* When, however, it is thrown upward, it goes more and more slowly; it has *negative acceleration* or *retardation.*

133. Relation of speed to time at constant acceleration. If we know the acceleration of any body, we can easily compute its speed at any time after it started.

For example, if the rate of acceleration of a train is 0.2 miles-per-hour per second, how fast is it moving one minute after it starts? One minute equals 60 seconds. If the train gains 0.2 miles-per-hour every second, then its speed, 60 seconds after starting, would be 60 times 0.2, or 12 miles per hour.

LAW I. *If the acceleration is constant, the speed acquired is directly proportional to the time.*

Final velocity = acceleration × time.

$$v = at. \qquad \text{(I)}$$

PROBLEMS

(Assume constant acceleration.)

1. Express 32 feet-per-second per second in miles-per-hour per second.

2. A body has a speed of 16 feet per second at a certain instant, and 3 seconds later it has a speed of 112 feet per second. What is its acceleration?

3. A train starting from rest has, after 33 seconds, a speed of 15 miles an hour. What is the average acceleration,

(*a*) In miles-per-hour per second?

(*b*) In feet-per-second per second?

4. If a locomotive can give a train an acceleration of 5 feet-per-second per second, how long will it take, after slowing down for a crossing, to increase the speed of the train from 22 feet per second to 82 feet per second?

5. What is the acceleration of a train if the initial speed is 45 feet per second, and after 5 seconds the speed is 15 feet per second?

6. The negative acceleration (retardation) in stopping electric trains is seldom greater than 4 feet-per-second per second. How long does it take to stop a train from 60 miles an hour?

7. Which of the following accelerations is the largest: —

(*a*) One foot-per-second per five seconds,

(*b*) One foot-per-five-seconds per second,

(*c*) One fifth of a foot-per-second per second?

134. Relation of distance to time at constant acceleration. Suppose a sled gains speed at a constant rate as it goes down a hill. If its acceleration is 3 feet-per-second per second, how far will it go in the first five seconds after starting from rest? We have already seen that its velocity at the end of five seconds will be 5×3, or 15 feet per second. Now it started from rest, that is, its *initial velocity* was zero, and gradually its speed increased until its *final velocity*, at the end of 5 seconds, is 15 feet per second. Therefore its *average velocity* is one half the sum of its initial and final velocities, or 7.5 feet per second.

$$\textbf{Average velocity} = \frac{\textbf{initial velocity + final velocity}}{2},$$

$$= \frac{v_0 + v_1}{2}.$$

We have already learned (section 129) that the distance traversed is the product of the average velocity and the time. So in this case the sled has gone 7.5×5, or 37.5 feet.

In general, for a body *starting from rest*, the average velocity is one half the final velocity;

$$\text{Average velocity} = \frac{v}{2}.$$

But we already know that the final velocity is $v = at$; then,

$$\text{Average velocity} = \frac{at}{2}.$$

Therefore the distance is

$$s = \frac{at}{2} \times t = \frac{1}{2} at^2. \qquad \text{(II)}$$

LAW II. *If the acceleration is constant, the distance traversed from rest varies as the square of the time.*

In using this law, acceleration should be expressed in ft./sec.^{2} or m./sec.^{2} or cm./sec.^{2}, and t in seconds.

135. Relation of speed to distance at constant acceleration. Suppose we wished to know how far the rapid transit electric car mentioned in the table in section 131 would have to run to develop a speed of 30 miles an hour, starting from rest. Since the question is concerned only with speed, distance, and acceleration, it is convenient to have an equation involving only v, s, and t.

From equation (I), we have

$$t = \frac{v}{a},$$

and from equation (II),

$$s = \frac{1}{2} at^2 = \frac{1}{2} a \times \frac{v^2}{a^2} = \frac{v^2}{2a}.$$

Then,

$$v^2 = 2as. \qquad \text{(III)}$$

LAW III. *If the acceleration is constant, the speed varies as the square root of the distance traversed.*

Equation (III) enables us to answer the question about the electric car.

$v = 30$ miles per hour $= 44$ feet per second.

$a =$ (say) 2.0 miles-per-hour per second $= 2.93$ feet-per-second per second.

$$s = \frac{v^2}{2a} = \frac{44^2}{2 \times 2.93} = 330 \text{ feet.}$$

Notice that 30 miles per hour and 2.0 miles-per-hour per second could not be substituted directly, because two different kinds of time units, namely hours and seconds, are involved. This is an example of the general rule that all the quantities substituted in any equation must first be expressed in **consistent units.**

It will save time to *memorize* equations (I), (II), and (III). Notice that there is an equation for each pair of quantities v and t, s and t, and v and s. *Always use the one equation that gives what is wanted directly from the data.*

136. Negative acceleration. Suppose that an engineer, running at 50 miles an hour, sees a child on the track 200 yards ahead. If his emergency air brakes can give him a retardation of 4 feet-per-second per second, can he stop in time?

Here we have a problem in retardation or negative acceleration. Let us think of the problem the other way around. Evidently if the engineer could stop within a given distance at a given retardation, he could get up speed within the same distance with an equally great acceleration. So we may ask instead, whether the engineer could get up to a speed of 50 miles an hour within 200 yards, if accelerating at 4 feet-per-second per second. The answers to the two questions are the same.

Since the quantities involved are a velocity v, and a distance s, we will use equation (III).

$v = 50$ miles an hour $= 73.3$ ft./sec.

$a = 4$ ft./sec.^{2}.

Then $$s = \frac{v^2}{2a} = \frac{(73.3)^2}{2 \times 4} = 672 \text{ feet} = 224 \text{ yards.}$$

So the engineer could *not* stop in time.

Problems

(Assume constant acceleration.)

1. If a locomotive can give its train an acceleration of 5 feet-per-second per second, in what distance can it develop a speed of 60 feet per second, starting from rest?

2. A boy runs toward an icy place in the sidewalk at a speed of 20 feet per second and slides on it 16 feet. What is the (negative) acceleration?

3. How far will a marble travel down an inclined plane in 3 seconds, if the acceleration is 50 centimeters-per-second per second?

4. A motor cycle starting from rest acquires a velocity of 40 miles an hour in 2 minutes. What is the acceleration in miles-per-hour per second?

5. How many yards does the motor cycle have to run in problem 4?

6. Two suburban stations are 2700 feet apart on a straight track. The greatest practicable acceleration or retardation is 3 feet-per-second per second. If there is no limit to the speed *en route*, what is the shortest possible running time between the stations, with stops at both?

137. Falling is motion at constant acceleration. It is possible to determine in the laboratory the time it takes a body to fall various distances. The results of an actual series of such experiments are as follows: —

Distances	Times	Ratio of Times
36 cm.	0.272 sec.	3
64	0.363	4
100	0.452	5
144	0.542	6

It will be seen that these distances vary almost exactly as the squares of the times, which we have seen to be the case when the acceleration is constant (see law II). Therefore falling is a case of motion at constant acceleration.

A freely falling body acquires velocity so rapidly that it is difficult to make observations upon it directly. Long ago

Galileo hit upon the plan of studying the laws of falling bodies by letting a ball roll down an incline. In this way he "diluted" the force of gravity and increased the time of fall so that it could be measured more accurately.

138. Galileo's experiment on the inclined plane. Galileo cut a trough one inch wide in a board 12 yards long, and rolled a brass ball down the trough. After about one hundred trials made for different inclinations and distances, he concluded that the distance of descent for a given inclination varied very nearly as the square of the time. It is remarkable that he was so successful in this experiment when we consider how he measured the time. He attached a very small spout to the bottom of a water pail and caught in a cup the water that escaped during the time the ball rolled down a given distance. Then the water was weighed and the times of descent were taken as proportional to the ascertained weight.

These experiments of Galileo are especially interesting because they led him to change his theories about the distance and time of falling bodies. He seems to have been one of the first of the ancient philosophers who thought it worth while to subject his theories to the test of experiment.

139. All freely falling bodies have the same acceleration. Before the time of Galileo (1564–1642) people believed that heavy objects fell faster than light objects; in other words, that the speed of a falling body depended upon its weight. But he claimed that all bodies, if unimpeded by the air, fell the same distance in the same time, and that the only thing that caused some objects, like pieces of paper or feathers, to fall more slowly than pieces of metal or coins, was the resistance of the air. To convince his doubting friends and associates he caused balls of different sizes and materials to be dropped at the same instant from the top of the leaning tower of Pisa. They saw the balls start together, and fall together, and heard them strike the ground together. Some

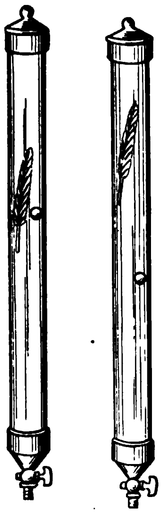


FIG. 134. — Feather and coin fall together in a vacuum.

were convinced, others returned to their rooms to consult the books of the old Greek philosopher, Aristotle, distrusting the evidence of their senses.

Later when the vacuum pump was invented, the truth of Galileo's view was confirmed by dropping a feather and a coin in a vacuum tube.

If we place a piece of metal and some light object, like a bit of paper or pith, or a feather, in a long tube (Fig. 134), and pump out the air, we find that, when we suddenly invert the tube, the two objects fall side by side from the top to the bottom. If we open the stopcock, letting the air in again, and repeat the experiment, we find that the metal falls to the bottom first.

140. Value of acceleration of gravity. It is possible to determine the value of the acceleration of gravity from the experimental data obtained in measuring the time of a free fall (section 137), and it is also possible to compute the value of this constant from the data got in the experiment of rolling a ball down an incline. Neither of these methods, however, yields as precise results as are obtained in experiments with pendulums.

We are all familiar with the pendulum as a means used to regulate the motion of clocks. It was long ago discovered by Galileo that the successive small vibrations of a pendulum are made in equal times, and that the time of vibration does not depend on the weight or nature of the bob, or the length of the swing, but does vary directly as the square root of the length of the pendulum, and inversely as the acceleration of gravity. This is expressed in the following formula: —

$$t = 2\pi\sqrt{\frac{l}{g}},$$

where t is the time in seconds of a complete vibration, l is the length of the pendulum in centimeters, g is the acceleration of gravity in centimeters-per-second per second, and π is 3.14.

We can measure t and l directly and π is known, so we may compute g from the formula; thus,

$$g = \frac{4\pi^2 l}{t^2}.$$

The value of the acceleration of gravity is about **980** *centimeters-per-second per second, or about* **32.2** *feet-per-second per second.* It varies a little from place to place.

Problems about falling bodies are just like other problems with constant acceleration. In the equations we usually represent the *acceleration of gravity* by g.

Thus, **for bodies falling freely from rest,**

$$\begin{aligned} v &= gt, \\ s &= \tfrac{1}{2} gt^2, \\ v^2 &= 2\,gs. \end{aligned}$$

It will be useful to remember that the speed with which a body must be projected upward to rise to a given height is the same as the velocity which it will acquire in falling from the same height. (Compare this statement with section 136.)

PROBLEMS

(Neglect air resistance.)

1. Make a table like the following, running up to $t = 5$ seconds, and fill it in.

NUMBER OF SECONDS, t	TOTAL DISTANCE FALLEN, s (FT.)	SPEED AT END OF EACH SECOND, v (FT./SEC.)	TOTAL DISTANCE FALLEN, s (METERS)	SPEED AT END OF EACH SECOND, v (M./SEC.)
1	16.1	32.2	4.9	9.8
2	?	?	?	?

2. A stone is dropped from the top of a cliff and strikes at the base in 5 seconds. (*a*) What velocity did it acquire? (*b*) How high is the cliff?

3. If a falling body has acquired a velocity of 150 feet per second, how long has it been falling? How far?

4. How many centimeters does a stone fall in 0.5 seconds?

5. How many centimeters does a stone fall during the fifth second?

6. A rifle is fired straight up (for speed, see table in section 129). How long before the bullet comes down again? How high will it go? (Assume that air resistance is negligible, which is far from true.)

7. A baseball is thrown up in the air and reaches the ground after 4 seconds. How high did it rise?

8. The weight of a pile driver drops 5 feet at first and later 15 feet. How much faster is it moving when it strikes in the latter case than in the first case?

9. A body is thrown vertically upward with a velocity of 50 meters per second. With what velocity will it pass a point 100 meters from the ground? (HINT. — How high does the body rise?)

10. How long would it take a bomb to fall 1000 feet from an aëroplane? During the fall the bomb would continue to move sidewise with the same velocity as the aëroplane, and so would always be directly under it. If the speed of the aëroplane is 60 miles an hour, how far will the bomb move sidewise while it is falling?

SUMMARY OF PRINCIPLES IN CHAPTER VII

$$\text{Speed} = \frac{\text{distance}}{\text{time}}.$$

$$\text{Acceleration} = \frac{\text{gain in speed}}{\text{time}}.$$

Laws of motion at constant acceleration: —

$$\text{I.} \quad v = at,$$

$$\text{II.} \quad s = \frac{at^2}{2},$$

$$\text{III.} \quad v^2 = 2\,as.$$

Value of acceleration of gravity: —

$$g = 32.2 \text{ ft./sec.}^2 = 980 \text{ cm./sec.}^2.$$

Questions

1. If you take two sheets of paper of the same size, and roll one of them into a ball, and let both the ball and the sheet of paper fall at the same instant from the same height, what is the result? Why?

2. How must the pendulum bob be moved on a clock which is running too fast?

3. What takes the place of a pendulum in a watch?

CHAPTER VIII

FORCE AND ACCELERATION

Inertia — the fundamental proportion — action and reaction — mass.

141. Newton's laws of motion. We are studying motion, and so far we have considered *how* bodies move; that is, we have been *describing* different motions, such as motion at constant speed and motion at constant acceleration. Now we shall begin to study *why* bodies move; we shall try to explain different motions by studying the forces that cause them. Practically all that we know about this part of physics dates back to Sir Isaac Newton (1642–1727), who wrote a treatise on the principles (*Principia*) of natural philosophy or physics. His whole book, and indeed all mechanics since his day, is based on three very simple laws, called **Newton's laws**. The **first** of them is the law of **inertia**, the **second** the law of **acceleration**, and the **third** the law of **interaction**. These will now be discussed in turn.

142. First law — Inertia. It is a familiar fact that nothing in nature will either start or stop moving of itself. Some force from outside is always required. For example, a horse when starting a wagon, even on an excellent road, has to pull very hard at first; once the wagon is going, the horse can keep it moving with very little effort; but if he tries to stop it to avoid running over some one, he has to push back hard. So also when a moving ship collides with another ship or a dock, it requires an enormous retarding force to stop her. In 1908 the *Florida* rammed the *Republic*, and her bow was crumpled back 30 feet before the force stopped her.

SIR ISAAC NEWTON. Born in England, in 1642. Died in 1727, and is buried in Westminster Abbey. Founded the science of mechanics, and made many important discoveries in light. Famous also for his achievements in mathematics and astronomy.

This inability of matter to change its state of motion (or of rest), except it be influenced from outside, is called **inertia**.

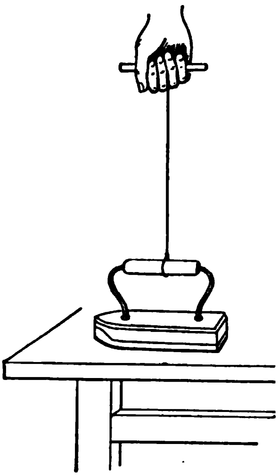


FIG. 136. — Inertia holds the weight still.

We may illustrate this property of inertia by balancing a card on a finger with a coin on top. Then we may snap the card out, leaving the coin on the finger. The coin moves only a little because there is only a small force due to friction to get it started. This may also be done with the apparatus shown in figure 135.

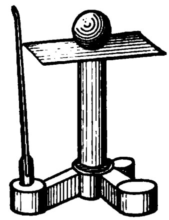


FIG. 135. — Inertia keeps the ball from moving.

Another interesting experiment is to try to pick up a flatiron by means of a linen thread tied to it (Fig. 136). If we pull slowly, we may be able to do this, but if we pull with a jerk, the string always breaks, because so much extra force is required to set the flatiron in motion quickly.

This familiar fact that bodies act as if disinclined to change their state, whether of rest or motion, was expressed by Newton in the following way: —

LAW I. *Every body persists in a state of rest, or of uniform motion in a straight line, unless compelled by external forces to change that state.*

QUESTION

If you roll a ball along the ground, it does not keep going indefinitely.
An automobile can start by itself.
Do these facts controvert Newton's First Law?

143. Applications of inertia. A nail can easily be driven into a heavy piece of wood, even when the wood does not lie on a firm foundation, because the quick blow of a hammer does not set the heavy piece of wood in motion to any great

extent. It is very difficult, however, to drive a nail through a light stick unless the stick is placed upon a solid foundation, or unless the stick is steadied by the inertia of a heavy sledge hammer, as shown in figure 137.

When the head of a hammer comes off, the best way to drive it on again is to hit the other end of the handle, rather than the head, against some solid foundation or with another hammer. (Why?)

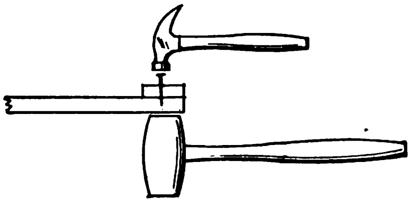


FIG. 137. — Inertia of sledge hammer.

144. Inertia in curved motion. This tendency of a body to continue to move in a straight line is very evident when it is desirable to make the body move in a circle. In this common case, a force is required to pull the body in toward the center of the circle, so that it may not fly off on a tangent. Such a force is called **a centripetal force,** meaning a force directed toward the center.

When an athlete swings a 16-pound hammer around his head before throwing it, he has to pull it inward because of its inertia. When he stops pulling inward, it flies off on a tangent. So all he has to do to throw it is to let go.

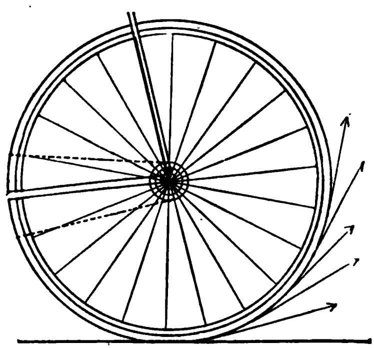


FIG. 138. — Mud flies off on a tangent.

Emery wheels revolve very rapidly. Sometimes one bursts because the cohesion between its parts is not enough to supply the centripetal force necessary to keep these various parts moving in their respective circles.

The mud on a bicycle wheel stays on the wheel only if the

adhesion between it and the tire is great enough to pull it around with the tire; otherwise it flies off on a tangent.

In a cream separator the denser part of the milk gets outside and crowds the lighter cream inward. This is because the greater inertia of the milk (that is, its greater tendency to move along a tangent) prevails over that of the cream.

When a train goes around a curve, the flanges of the wheels are pressed inward by the outer rail; if the rail is not strong enough to exert the necessary force inward, the train is wrecked on the outside of the roadbed. This is made clear in figure 139. The *weight* of the train is balanced by the *upward* push A of the tracks, the *centripetal* force B is exerted *inward* by the rails against the flanges, and therefore the resultant R is inclined. Consequently the track should be tilted toward the center, that is, "banked," so as to be at right angles to R. This equalizes the pressure on both sides and relieves the pressure on the outside flanges, thus making them less likely to break.

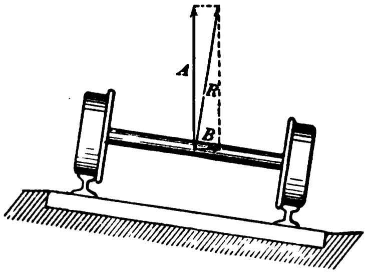


FIG. 139. — Banking rails on a curve.

145. Second law — Acceleration. We have been discussing what happens to a body when forces *do not* act on it. Let us now consider what happens when forces *do* act on it.

Whenever an "unbalanced" force is acting on a body, the body has an acceleration in the direction in which the force acts, and the acceleration is proportional to the force. By an **unbalanced force** we mean more push or pull in one direction than in the other. For example, a locomotive is pulling a train at a constant speed of 50 miles an hour. The engine is certainly exerting a force on the train, but there are other forces, due to friction and air resistance, acting in the opposite direction, and these just balance the pull of the

engine. The *net* force forward is zero; if it was not zero, the train would not only be *going* forward but *accelerating* forward; it would be *gaining speed.*

It is important to keep in mind that it is *net force* and *acceleration* which always go together, and not net force and motion. The above example shows that we can have motion without net force if the speed is not changing.

LAW II. *The acceleration of a given body is proportional to the force causing it.*

That is, if any given body is acted on at one time by a force F_1, and at another time by another force F_2, then

$$\frac{F_1}{F_2}=\frac{a_1}{a_2},$$

where a_1 and a_2 are the accelerations produced by F_1 and F_2.

In other words, if we pull a body with a certain force, and at another time pull it twice as hard, it will have twice as much acceleration the second time as the first.

One way to cause a force to act on a body is to let the body fall. In this case the force acting is known, namely, the weight W of the body. The acceleration is also known, namely, g, which is 32.2 feet-per-second per second, or 980 centimeters-per-second per second. So the weight of the body and its acceleration when falling can always be used as two of the numbers in a proportion.

That is, $$\frac{F}{W}=\frac{a}{g}.$$

This enables us to compute the force needed to give a certain body any desired acceleration.

For example, a freight train weighs 1000 tons. How great a force is necessary to give it an acceleration of half a foot-per-second per second?

$$\frac{F}{1000}=\frac{0.5}{32.2},$$

$$F=\frac{1000\times 0.5}{32.2}=15.6 \text{ tons.}$$

146. Units. In the equation $F/W = a/g$, it makes no difference in what unit F and W are expressed, *provided only that both are expressed in the same unit.* Both can be expressed in pounds, or in ounces, or in tons, or in kilograms, or in grams, or in a less familiar unit called a "dyne." The **dyne** is a very small unit of force much used in scientific work, especially electrical measurements. It can be defined as 1/980 of a gram weight.* It is about the weight of a milligram. If a force is given in terms of any one of these units, it can be expressed in terms of any other of them with the help of the following table: —

1 gram = 980 dynes.	1 dyne = 0.00102 grams.
1 pound = 454 grams.	1 gram = 0.00220 pounds.
1 pound = 445,000 dynes.	1 dyne = 0.00000225 pounds.

Similarly a and g may be in any units, *provided only that both are in the same unit.* If both are to be in feet-per-second per second, the numerical value of g is 32.2; if both are to be in centimeters-per-second per second, the numerical value of g is 980.

Problems

1. Express 110 grams in pounds.
2. Express 110 grams in dynes.
3. Express 8,000,000 dynes in pounds.
4. What acceleration will a force of 5 pounds produce in a body weighing 16.1 pounds?
5. What acceleration will a force of 1 gram produce in a body weighing 327 grams?
6. What acceleration will a force of 1 pound produce in a body weighing 1 pound?
7. What acceleration will a force of 1 dyne produce in a body weighing 1 gram? (Note. — The answer to this problem is often regarded as the definition of a dyne.)
8. State accurately in words the definition of a dyne that is referred to in the last problem.

* See also problems 7 and 8 below.

9. A body weighing 10 pounds is observed to have an acceleration of 2 feet-per-second per second. What force is acting?

10. A force of 1 kilogram is observed to produce an acceleration of 9.8 centimeters-per-second per second in a certain body. How much does the body weigh?

11. A force of 1000 dynes is observed to produce an acceleration of 9.8 centimeters-per-second per second in a certain body. How many grams does the body weigh?

12. An automobile weighing 2 tons is started from rest with an acceleration of 4 feet-per-second per second. How hard is the road pushing the bottoms of the rear tires forward?

13. An elevator weighing 980 kilograms is pulled upward by a force great enough to hold up the weight and give 200 kilograms of *net* force besides. What is the acceleration of the elevator?

14. What pressure will a 150-pound man exert on the floor of an elevator which is going up with an acceleration of 4 feet-per-second per second?

15. A train starting from rest with a constant acceleration takes 44 seconds to get up to a speed of 30 miles an hour. If the train consists of 4 all-steel cars, each weighing with its load 62.5 tons, what pull is exerted by the engine? (HINT. — Compute acceleration and then find force.)

147. Third law — Interaction. Newton's third law is based on two familiar facts. One way of stating the first of these facts is that there can never be a force acting in nature unless *two* bodies are involved, one exerting it and one on which it is exerted. Thus, when a railroad train is pulled, there is an engine that does the pulling; and on the other hand, the engine cannot exert a pull or a push without something to be pulled or pushed. An electric car or an automobile seems, perhaps, to push itself along, but really the track or the road under the wheels is exerting a force on the wheels and pushing the car along. We have all seen what happens when the car track is so icy or the road so muddy that it cannot push on the wheels. The motor is going just as hard as ever, but the car does not move.

In order to make this idea seem more real to us, let us try the experiment on a small scale, as shown in figure 140. If we wind up the little toy engine, and place it on the circular track, which is so mounted as to turn

easily, we find that the track turns around and the rails under the wheels go backwards. If we hold the track fast, the engine goes ahead twice as fast as at first, and if we hold the engine fast, the track turns around backwards twice as fast as at first.

Another case is that of any heavy object: there is a force called its weight (force of gravity) pulling it down; but we know that it is the earth that exerts this force.

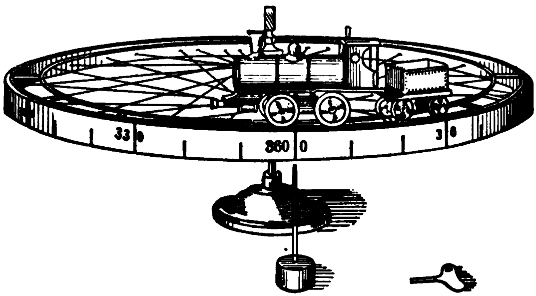


FIG. 140. — Track pushes the engine forward.

This, then, is the *first* fact: whenever there is a force in nature there must be *two* bodies, one to exert it and one to receive it.

But we can go farther than this. We can say that whenever there is a force in nature, there must be not only two bodies involved, but *another force*. That is, forces never exist singly, but always in pairs. If the first force was exerted by a locomotive on a train, the second will be exerted by the train on the locomotive. The train will pull back on the locomotive just as hard as the locomotive pulls forward on the train. If a road is pushing forward on the wheels of the automobile, the wheels must be pushing back on the road.

If, instead of the road, we substitute great rollers, we may measure this backward push. This is the method sometimes used in testing laboratories in making power tests of automobiles and locomotives.

Finally, when any heavy object is pulled downward by the earth, the heavy object must be pulling the earth up with an equal force. This does not seem very likely at first, but this is simply because the force is so small and the earth so large that the force has an imperceptible effect on the earth. If the heavy body which we are thinking of is the moon, the whole thing becomes reasonable at once, for the earth and the moon are actually rotating about a point O (Fig. 141), which is not exactly at the center of the earth. So the moon must continually pull the earth to make its center of gravity move in its circle.

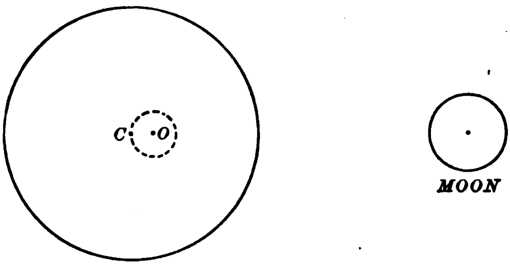


FIG. 141. — Rotation of the moon about the earth.

This fact that forces always occur in pairs, one of the pair being equal and opposite to the other, was expressed by Newton in the following form: —

LAW III. *With every action (or force) there is an equal and opposite reaction.*

148. Mass *vs.* weight. "Mass" and "weight" are constantly confused in ordinary conversation. While we have preferred not to use the term "mass" in studying Newton's second law, yet it is well to know its precise meaning that we may read intelligently the books which make use of it.

Mass means quantity of matter. It is the answer to the question, "How much matter is there in a given body?"

Weight means the pull of gravity on the body. The weight of a body is a *force* acting on the body, not a description of what it contains.

The unit of mass is the quantity of matter contained in a certain piece of platinum (the standard kilogram, Fig. 142).

The unit of weight is the pull of the earth on that same piece of platinum, when it is near sea level and at latitude 45°.

Since a kilogram mass weighs a kilogram under these standard conditions, the mass and the "standard weight" of a body are numerically equal. But if we carry a kilogram mass to the top of a high mountain, and weigh it on a very sensitive *spring* balance, it will weigh less than a kilogram, because it is farther from the center of the earth, and so the earth pulls less hard on it. The reading of the spring balance might be called its "local weight."

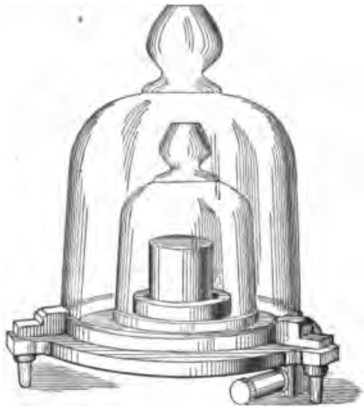


FIG. 142. — Standard kilogram.

Since all bodies on the mountain top would weigh less in the same proportion, we can get the *standard* weight of anything without descending the mountain by weighing it on an *equal-arm* balance against a set of "standard weights." This is what we always do in the laboratory and in the outside world when we want to know weights accurately. So when we speak of the weight of a body we almost always mean its "standard weight."

Since $F = \frac{W}{g}a$, and since the standard weight W and the mass M are numerically equal, we shall get the *same value* for F if we write (when using grams, centimeters, and seconds)

$$F = \frac{Ma}{980},$$

or

$$980\,F = Ma.$$

Here F is in *grams;* if we choose, however, to express the force as F' dynes, instead of as F grams, then F and F' will be different numbers, and

$$F' = 980\ F,$$

so $$F' \text{(dynes)} = M\text{(grams)} \times a\,(\text{cm./sec.}^2).$$

This is a common way of expressing Newton's second law.

SUMMARY OF PRINCIPLES IN CHAPTER VIII

Newton's laws and the fundamental proportion: —

I. Every body continues in a state of rest or of uniform motion in a straight line, unless compelled by external forces to change that state.

II. The acceleration of a given body is proportional to the force causing it.

$$\frac{F}{W} = \frac{a}{g}.$$

III. With every action (or force) there is an equal and opposite reaction.

Distinction between mass and weight.

QUESTIONS

1. How is the water quickly removed from wet clothes in a steam laundry?
2. Why does a train continue to move after the steam is shut off?
3. Why do automobiles "skid" in rounding corners rapidly?
4. What does an aviator have to do to round a corner safely, and why?
5. Why can small emery wheels be safely driven at a greater speed, that is, at more revolutions per minute, than larger ones?
6. Why does a wheel, or any revolving part of a machine, sometimes shake or hammer in its bearings?
7. Explain how a locomotive engineer can tell, when he starts up his train, if one of the cars has been uncoupled from the train.
8. Explain why lawn sprinklers rotate. Would such a sprinkler rotate in a vacuum?

CHAPTER IX

ENERGY AND MOMENTUM

Kinetic energy — the law of energy — potential energy — the conservation of energy — momentum and its law.

149. Kinetic energy. We have already seen (section 31) that in physics *work* involves not only *force* but also *displacement.* Whenever a force moves anything in its own direction, the force does work *on* the thing, and whenever anything moves against a force, the thing does work *against* the force.

The **energy** of anything may be defined as its **capacity for doing work.** Thus a heavy flywheel will keep machinery running for some time after the power has been shut off. Therefore, a heavy flywheel *in motion* can do work; it has *energy.* The energy of any body which is due to its **motion** is called **kinetic energy.**

Let us consider more carefully the case of a heavy flywheel on an engine. At first the engine has to push and pull on the crank shaft to get the flywheel started and to bring it up to speed; the engine has to do work on the flywheel. When once the flywheel is up to speed, however, the engine does not have to push or pull any longer to keep the flywheel going (except for friction, which we will neglect for the moment). From this time on all the work that the engine does goes into the driven machinery attached to the shaft. Suppose now that the pressure of the steam on the engine suddenly drops, or that an extra load is thrown on the shaft. The shaft does not stop turning suddenly, or drop instantly to a lower speed. It slows down gradually, pulling back on

the flywheel as it does so, and taking work out of the flywheel, that is, making the flywheel *do* work instead of *absorbing* it. This continues until the engine picks up the load again, or until the flywheel stops.

In other words, the flywheel, as long as it is moving, can do work on the shaft if necessary, and the faster it is moving, the more work it can do before it comes to rest. In physics we describe this very familiar fact by saying that the flywheel has **energy**, and since its energy depends upon its being in motion, we call it **kinetic energy**, or energy of motion.

Every body in motion has kinetic energy; that is, it will do a certain amount of work against a resisting force before it will stop. Furthermore the kinetic energy of a body will be greater the heavier it is and the faster it is moving.

150. How to measure kinetic energy. It will be easier to do this for a body moving straight ahead rather than in a circle. We will consider first how much work it takes to get a heavy body up to a given speed, and then how much work it will do before it stops. By definition this latter is its kinetic energy.

The force necessary to get a body started with a given acceleration a is, by the fundamental proportion (section 145),

$$F = \frac{W}{g}a.$$

If the distance which the body moves before it gets up to a given speed is called s, the work done is the product of the force by the distance, namely,

$$Fs = \frac{W}{g}as.$$

But it will be seen that the product as can be expressed in terms of the speed acquired, v, by means of the third law of accelerated motion (section 135). Thus,

$$v^2 = 2as,$$

or $$as = \frac{v^2}{2}.$$

So the work done is

$$Fs = \frac{Wv^2}{2g}.$$

Thus we see that the work required to bring a heavy body from rest up to a given speed does not depend on the acceleration, or on the distance covered while coming up to speed, but only on the weight of the body and the speed itself.

Now, how much work will the body do against a retarding force before it comes to rest? We have already seen that the easiest way to think of a retardation problem is as an acceleration problem reversed. That is, a body will stop under a given retarding force in the same distance that it would need to get up speed under an equal accelerating force, and it will do the same work against the retarding force that an equal accelerating force would have to do on it to get it started. So the formula above gives not only the work necessary to start it, but also the work it will do when it stops.

Therefore,

$$\textbf{Kinetic energy} = \frac{Wv^2}{2g}.$$

151. The energy equation. The equation just found, namely,

$$Fs = \frac{Wv^2}{2g},$$

is called the energy equation. It applies, as we have just seen, either to accelerating or to retarding bodies, if they start from or come to rest. It can be stated in words as follows: —

If the body is *gaining* speed,

Gain of kinetic energy = accelerating force × distance
= work done by force on body.

If the body is *losing* speed,

Loss of kinetic energy = retarding force × distance
= work done by body against force.

152. Units. In using this equation we must be consistent in our units. Thus F and W are both forces and both must be expressed in the same unit in any one application of the equation. In one problem we may choose tons and in another dynes, but in any single problem *all* forces must be in tons if we have chosen tons, and in dynes if we have chosen dynes.

In the same way, s, v, and g must all involve the same unit of length. In one problem we may choose centimeters and in another feet, but once we have started the problem we must stick to our choice.

In expressing the velocity, v, and the acceleration, g, it is customary always to use the *second* as the unit of time. Therefore g will always be either 32.2 ft./sec.^{2} or 980 cm./sec.^{2} according as we have chosen feet or centimeters as the unit of length.

The left-hand side of the equation, Fs, is force times distance, or work, and so the right-hand member, which is equal to it, will come out expressed in *work units*. There are several **work units** in common use, such as the

foot pound (ft. lb.),
foot ton (ft. T.),
gram centimeter (g. cm.),
kilogram meter (kg. m.), and
dyne centimeter ("erg").

Since each of these work units is a force unit times a distance unit, we can always tell what unit the kinetic energy will come out in, if we notice what force unit and what distance unit we started with.

For example, if W is in pounds, v in feet per second, and g in feet-per-second per second ($g = 32.2$ ft./sec.^{2}), the kinetic energy ($Wv^2/2\,g$) will

be in foot pounds. But if W is expressed in dynes, v in centimeters per second, and g in centimeters-per-second per second ($g = 980$ cm./sec.^{2}), the kinetic energy ($Wv^2/2g$) will be in dyne centimeters. There is a shorter name for a "dyne centimeter"; it is usually called an **erg**. Since the erg is a very small unit of work, the **joule** $= 10^7$ ergs is often used.

153. Applications of the energy equation. The **energy equation** will help us to solve many useful problems about moving things which involve the question "how far," or the idea of **distance** in general.

For example, consider again the problem of the engineer and the child (section 136). Suppose that the train is going 50 miles an hour, but that we do not know its rate of retardation. If the retarding *force* is equal to one eighth of the weight of the train, how far will the train run before coming to a standstill?

We can compute the rate of retardation from the fundamental proportion, and then proceed as before, but it will be easier to use the energy equation as follows: —

The speed 50 miles an hour = 73.3 ft./sec.

So the kinetic energy is

$$\frac{W \times (73.3)^2}{2 \times 32.2} \text{ ft. lbs.}$$

The retarding force is $W/8$ lbs.

Therefore,
$$\frac{W}{8} \times s = \frac{W \times (73.3)^2}{2 \times 32.2},$$

and since the W's cancel out, this can be solved for s, giving

$$s = 667 \text{ feet} = 222 \text{ yards.}$$

So in this case also, the engineer could *not* stop in time.

Again, suppose a car weighing 10 tons is going 36 miles an hour. What force is required to stop it within a space of 100 feet?

The velocity must be expressed in feet per second since we ordinarily do not use g in miles/hour^2.

$$v = 36 \text{ miles/hour} = 52.8 \text{ ft./sec.}$$

But W and F can be left in tons.

The kinetic energy is $\dfrac{10 \times (52.8)^2}{2 \times 32.2} = 433$ ft. T.

So
$$F \times 100 = 433, \text{ or } F = 4.33 \text{ tons.}$$

Finally, suppose that the flywheel mentioned in section 149 has a 10 ton rim, and that we can neglect the effect of the thin spokes. Suppose also that it is 16 feet in diameter and making 15 revolutions per minute (r. p. m.). How much kinetic energy has it?

Each part of the rim is making 15 turns per minute and therefore moving with a velocity of $15 \times 2\ \pi r = 15 \times 2 \times 3.14 \times 8 = 754$ ft./min., which is equal to 12.5 ft./sec. Therefore, the kinetic energy of the whole rim is

$$\frac{10 \times (12.5)^2}{2 \times 32.2} = 24.3 \text{ foot tons.}$$

This is the same as 48,600 foot pounds. It is the amount of work which the flywheel could do before stopping.

Problems

(State the unit in which each answer is expressed.)

1. What is the kinetic energy of a baseball weighing one third of a pound if its velocity is 64.4 feet per second?
2. What is the kinetic energy of an 80-ton locomotive going 60 miles an hour?
3. What is the kinetic energy of a 9.8-kilogram weight which has been falling long enough to have a velocity of 12 meters per second?
4. What is the kinetic energy of a 16.1-gram bullet whose velocity is 600 meters per second?
5. Find the kinetic energy *in ergs* of a stone weighing 20 grams when it is thrown with a velocity of 800 centimeters per second.
6. The 14-inch guns on some of the United States warships fire a projectile weighing 1400 pounds and are said to give it a "muzzle energy" of 65,600 foot tons. What is the velocity of the projectile as it leaves the gun?
7. What resistance is necessary to stop a body whose kinetic energy is 90,000 ergs, in a distance of 3 meters?
8. A boy weighing 100 pounds starts to slide on ice at a speed of 20 feet per second. What is his initial kinetic energy? If the retarding force due to friction is 40 pounds, how far will he go before stopping?
9. How great a force in excess of that required to overcome friction is necessary to bring a 3220-pound automobile up to a speed of 30 miles an hour in a distance of 242 feet?

154. Potential energy. Some things have a capacity for doing work even when they are not in motion. Thus when a clock spring is wound up, it can drive the clock as it un-

coils, because of the elastic strain in it, due to its change of shape. If the clock has a weight instead of a spring, the weight can drive the clock because of its elevated position. Such energy, due to strain or to position, is called **potential energy**.

Just as the kinetic energy of a moving weight can be measured either by the work required to get it up to speed or by the work it will do while stopping, so the potential energy of a raised weight or of a coiled spring can be measured either by the work required to raise or coil it, or by the work it will do when it falls or unwinds.

In a later chapter we shall see that when a lump of coal burns, it gives out energy in another form called heat, some of which can be used to drive a steam engine. Thus the unburned coal has in it capacity to do work, that is, energy, and since this energy is not due to any motion of the lump of coal, it must also be potential. This kind of potential energy is usually called **chemical energy**.

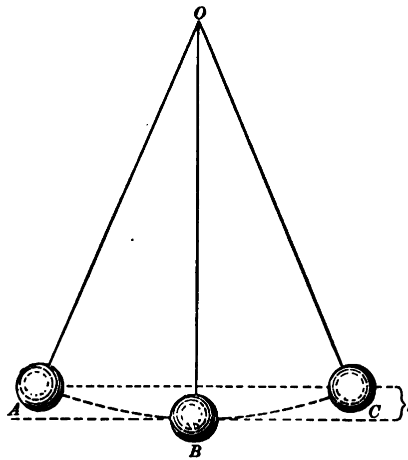


FIG. 143. — Transformation of energy in pendulum.

155. Transformation of energy. In nature the various forms of energy, kinetic or potential, are continually changing into one another.

For example, when a pendulum bob (Fig. 143) is at the highest part of its swing *A*, it has potential energy because of its height. As it swings down this potential energy disappears, but the bob gains speed and kinetic energy. As the bob swings up again on the other side, *C*, its velocity and kinetic energy decrease, but its potential energy increases.

Similarly when coal is burned, its chemical energy changes into heat. Some of this heat may be changed into potential energy in the form of steam under pressure. A steam engine could then change some of this into kinetic energy in a flywheel, or into some other form of mechanical energy in a driven machine, or into electrical energy in a dynamo. Some of this latter might be changed into light in a lamp, while the rest would turn back to heat.

In all these cases we may think of energy as flowing about from one place to another, passing through the various machines and having its outward appearance changed by them, almost like water flowing from a reservoir through a dyehouse, where it is used for many purposes, only finally to be dumped into a stream or sewer, changed in appearance, but unmistakably the same kind of thing that went in.

156. The conservation of energy. After its use in the dyehouse some of the water might never get through to the stream, having been used up in some chemical process, so that it is no longer water. But in the case of energy this cannot happen. Energy is *never made from* anything that is not energy, or *turned into* anything that is not energy. The total quantity of energy in the universe is always the same and is changed only in form and distribution. In any given machine there may be leaks of energy because of friction, radiation, etc., just as there may be leaks in the pipes in the dyehouse, but the energy that leaks away is not destroyed, but is given as heat to the surroundings of the machine, where it is of no more use than water spilled on the floor.

Thus in the pendulum (Fig. 143) the sum of the kinetic and potential energies is the same wherever it is in its swing, unless there is friction. If there is friction, some energy disappears as heat and less is left in the pendulum, but the total quantity, counting in the heat, is unchanged.

This fact, that *energy can never be manufactured or destroyed, but only transformed, and directed in its flow*, was first stated (although not very clearly) by a German, Robert

Mayer, in 1842. It is called the **law of the conservation of energy.** It has become the most important generalization in all physics, and its value will be more and more evident as we study the subject.

157. "Perpetual motion" machines. One of the most interesting applications of this principle is that it assures us that "perpetual motion" machines are impossible. Such a machine would be one that runs of itself, without being driven by an engine, and without burning any fuel, and does something useful without cost. Such a machine, if it could be built, would be of extraordinary value to its inventor and to the world, and for hundreds of years people have been trying to invent one. But the principle of the conservation of energy shows that no such machine can possibly be made, because it would be manufacturing energy out of nothing.

158. Momentum and energy. In the colloquial use of these words there is a great deal of confusion. When a person is thinking of something which a body does because it is moving, he is likely to talk about either its "momentum" or its "energy," whichever word first occurs to him. It is therefore worth while to take a little trouble to understand clearly the difference between momentum and energy.

We have seen that when a force acts on a moving body through a long *distance*, it accomplishes more than when the distance is short. The *work* done is greater. It is also evident that when a force acts on a moving body for a long *time*, it accomplishes more than when the time is short. In the technical language of physics we say that the **impulse** of the force is greater. *Impulse may be defined as the force multiplied by the length of time it acts.* Thus,

Work = force × distance,
Impulse = force × time.

We shall find that **momentum** has the same relation to **impulse** that kinetic energy has to work. The idea of momentum will help us to solve problems involving "how long?" just as the idea of kinetic energy helps us to solve

problems involving "how far?" To show this we will print again the proof of the energy equation side by side with the corresponding proof of a momentum equation.

PROOF OF ENERGY EQUATION		PROOF OF MOMENTUM EQUATION	
	$F = \frac{W}{g} a.$		$F = \frac{W}{g} a.$
	$Fs = \frac{W}{g} as.$		$Ft = \frac{W}{g} at.$
But	$v^2 = 2\,as$ or $\frac{v^2}{2} = as.$	But	$v = at.$
So	$\boldsymbol{Fs = \frac{Wv^2}{2\,g}}.$	So	$\boldsymbol{Ft = \frac{Wv}{g}}.$

The expression $\frac{Wv}{g}$ *is called the momentum of a moving body.*

Both kinetic energy and momentum are proportional to the *weight* of the moving body; thus a railroad train has both more momentum and more energy than a motor cycle running at the same speed.

In the second place both the momentum and the energy of a moving body increase when its *speed* increases, but not according to the same law. The *energy* is proportional to the *square of the speed.* That is, doubling the speed of a train makes its kinetic energy *four* times as large. But the momentum is proportional only to the *first power* of the speed. That is, doubling the speed of a train merely *doubles* its momentum.

Finally, we must not forget that there is a *two* (2) in the denominator of the expression for energy, but not in the expression for momentum.

159. The momentum equation. The equation

$$\boldsymbol{Ft = \frac{Wv}{g}}$$

is called the momentum equation. It holds either for accelerating or retarding bodies, and can be expressed in words as follows:—

If the body is *gaining* speed,

Gain of momentum = accelerating force × time
= impulse received from force.

If the body is *losing* speed,

Loss of momentum = retarding force × time
= impulse lost to force.

160. Units of momentum. In using the momentum equation, as in using the energy equation, we must be consistent in our units. That is, F and W must be in the same unit of force, and v and g must involve the same unit of length. Furthermore, v, g, and t must all be expressed in terms of seconds, because it is not worth while to remember any other way of expressing g than 32.2 ft./sec.^{2} or 980 cm./sec.^{2}.

The momentum equation shows that a momentum Wv/g is equal to a *force* times a *time*, and so the unit of momentum will be a force unit times a time unit. Thus, momentum may be expressed as

pound seconds, or
ton seconds, or
gram seconds, or
kilogram seconds, or
dyne seconds,

according to the *force* and *time* units used in the equation.

The *dyne second* as a unit of momentum corresponds to the *dyne centimeter* or *erg* as a unit of energy, but curiously no one has ever thought it worth while to give it a name of its own corresponding to "erg."

161. Applications of the momentum equation. The momentum equation will help us to solve many problems about moving things which involve the question "how long?" or the idea of **time** in general.

For example, a certain engine can exert a "drawbar" pull on its train equal to $\frac{1}{40}$ of the weight of the train. How long will it take to bring the train up to a speed of 50 miles an hour, starting from rest?

$$v = 50 \text{ miles/hour} = 73.3 \text{ ft./sec.}$$

$$\frac{W}{40} \times t = \frac{W \times 73.3}{32.2},$$

and since the W's cancel out,

$$t = 91 \text{ seconds.}$$

Again, an electric car weighing 12 tons, running 15 miles an hour, can stop in 7 seconds. What is the retarding force?

$$v = 15 \text{ miles hour} = 22 \text{ ft./sec.}$$

$$F \times 7 = \frac{12 \times 22}{32.2} \text{ ton seconds.}$$

$$F = 1.17 \text{ tons.}$$

Again, a steamboat weighing 20,000 tons is being pulled by a tugboat, which exerts a force great enough to overcome the friction of the water and to give a *net* force of 2 tons besides. What speed will the boat acquire in 4 minutes, starting from rest?

$$2 \times 4 \times 60 = \frac{20000 \times v}{32.2},$$

$$v = 0.773 \text{ ft./sec.}$$

Another application, which is important in studying steam turbines, windmills, and aëroplanes, is the case of a steady stream of fluid striking against a solid surface. For this purpose we may write the equation as

$$F = \frac{W}{t} \times \frac{v}{g},$$

and W/t is then the weight of fluid striking the surface *per second*.

One of the useful applications of the momentum equation is in studying a blow, such as a bat gives a ball, or a collision, as between two billiard balls. We naturally speak of the forces acting in such cases as "impulses," and this explains why the product $F \times t$, which was first used in solving such problems, is called "impulse."

Problems

(State the unit in which each answer is expressed.)

1. What is the momentum of a 180-pound football player running at a speed of 20 feet per second?

2. What is the momentum of a 66,000-ton ship when it is going 24 miles an hour (about 21 knots)?

3. How fast must a 1000-kilogram automobile be going to have 3000 kilogram seconds of momentum?

4. A car weighing 12 tons, moving 5 feet per second, is stopped by a bumper in 0.2 seconds. What is the average force of the blow?

5. An 8000-ton ship moving 4 miles an hour is stopped in 2 minutes. Find the average force.

6. A 5-pound hammer moving 40 feet per second strikes a nail. If the average resistance of the wood to the nail is 1240 pounds, what fraction of a second was required to bring the hammer to rest?

7. A fire engine throws a 2-inch stream of water horizontally against a brick wall with a velocity of 150 feet per second. What is the force exerted on the wall?

8. A gun delivers 100 bullets per minute, each weighing one ounce, with a horizontal velocity of 1500 feet per second. What is the average force exerted by the gun?

SUMMARY OF PRINCIPLES IN CHAPTER IX

The law of energy ("How far?").

Work = force × distance.

Kinetic energy $= \dfrac{Wv^2}{2g}$.

Work done on body = gain of kinetic energy.

Work done by body = loss of kinetic energy.

The conservation of energy: Energy can never be manufactured or destroyed, but only transformed or directed in its flow.

The law of momentum ("How long?").

Impulse = force × time.

Momentum $= \dfrac{Wv}{g}$.

Impulse given to body = gain of momentum.

Impulse given by body = loss of momentum.

QUESTIONS

1. The pendulum of a clock would die down because of the friction of the air around it if energy were not continually supplied to it. How is this done?

2. Look up "perpetual motion" machines in an encyclopedia and try to see for yourself why some of them cannot work.

3. A certain rifle was once described in the headline of a magazine advertisement as striking "a blow of 2038 pounds." Farther down in the advertisement it appeared that the bullet weighed $\frac{1}{35}$ of a pound, and that its velocity was 2142 feet per second. What did the headline mean?

CHAPTER X

HEAT — EXPANSION AND TRANSMISSION

Thermometer scales — linear and volumetric expansion of solids — expansion of liquids — maximum density of water — expansion of gases — pressure coefficient of gases — the gas thermometer and the absolute scale — gas formula — hot-air engine — convection currents — heat transfer by convection — heating and ventilation systems — conduction — radiation — molecular theory.

EXPANSION BY HEAT

162. Sources of heat. Our most important source of heat is the **sun.** The sun's rays give more heat, the more nearly vertical they are. This explains why we receive more heat at noon than in the morning or evening, and more heat in summer than in winter.

The **interior of the earth** also is hot. In mine shafts sunk into the earth the temperature rises about one degree for every hundred feet of depth. Hot springs and volcanoes also lead us to think that the inside of the earth is hot.

To warm our houses and run our engines, we do not as yet depend directly on the sun or on the heat in the earth, but on the heat produced in **burning** wood, coal, oil, or gas. The heat thus obtained comes indirectly from the sun, having been stored as chemical energy in plants in past ages.

We have already learned in our study of machines and in our everyday experience that **friction** produces heat. For example, in scratching a match, in using drills, saws, and files, indeed, whenever mechanical energy is apparently lost, we find that heat appears.

John Tyndall (1820–1893) in his lectures on "Heat considered as a mode of motion" used to perform a striking experiment to show that friction produces heat.

Let us try the same experiment by putting a little water in a metal tube (Fig. 144). If we close the tube with a stopper and rotate it either by hand or with a motor, we shall find that the friction between the rotating tube and the wooden clamp will generate in a few minutes enough heat to boil the water and blow the stopper out.

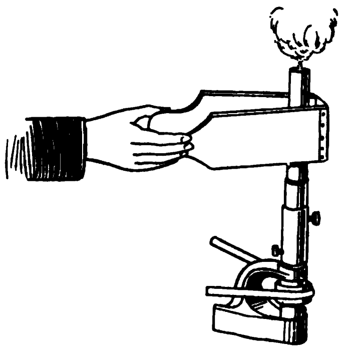


Fig. 144. — Boiling water by friction.

163. The thermometer. A deep cellar seems cold in summer and warm in winter, even though it remains at nearly the same temperature. A room often seems hot after we have been out in the cold, although it seems chilly after we have been in it awhile. Our sensations about the temperature of things are therefore very unreliable and depend on our own condition at the moment. So it is necessary to have some kind of instrument to indicate accurately how hot or cold things are, that is, a **thermometer.** The usual form of thermometer is based on the fact that most liquids, such as mercury and alcohol, expand when being heated and contract again on cooling.

164. Making a mercury thermometer. A spherical or cylindrical bulb is blown on one end of a piece of glass tubing with a very fine uniform bore, and the bulb and part of the stem are filled with mercury. When the mercury is warmed, it expands and rises in the stem until it overflows. Then the top of the tube is closed by melting the glass. When the mercury cools again, it leaves a vacuum in the top of the tube. If the bulb is now placed in the steam from boiling water, the mercury rises to a definite point on the

stem, which is marked with a scratch. This point is called the **boiling point.** If the thermometer is then put in melting ice, the mercury goes back down the stem and stops at a definite point. This point is called the **freezing point.**

In thermometers that are used for scientific work the distance on the stem between these two fixed points is divided into 100 equal spaces, called degrees. In this thermometer, which is called a **Centigrade** thermometer, the **freezing point** is marked **zero** and the **boiling point** is marked **one hundred.** When these divisions extend below the zero point, they are called degrees below zero or minus degrees.

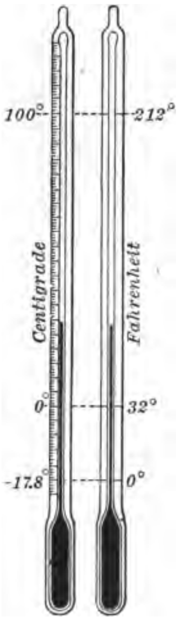


FIG. 145.—Centigrade and Fahrenheit scales.

165. Centigrade and Fahrenheit scales. In England and in North America a scale devised by **Fahrenheit** is in common use. On this scale the freezing point is marked 32 degrees (32°) and the boiling point 212°, so that the space between the freezing and boiling points is divided into 180 divisions (Fig. 145). Since 100 divisions on the Centigrade scale are equivalent to 180 divisions on the Fahrenheit scale, one division Centigrade is equivalent to $\frac{9}{5}$ divisions Fahrenheit. To change a temperature expressed on the Centigrade scale to the Fahrenheit scale, we have only to multiply by $\frac{9}{5}$ and add 32°. For example: —

$$30^\circ \text{ C} = (\tfrac{9}{5} \times 30) + 32 = 86^\circ \text{ F}.$$

To change a temperature expressed on the Fahrenheit scale to the Centigrade scale, we must first subtract 32° and then multiply by $\frac{5}{9}$. For example: —

$$98.6^\circ \text{ F} = (98.6 - 32)\tfrac{5}{9} = 37^\circ \text{ C}.$$

Inasmuch as mercury freezes at −39° C, the thermometers used for very low temperatures contain alcohol, which is usually colored red or blue.

166. Special thermometers. In Weather Bureau stations the lowest temperature during the night and the highest temperature during the day are automatically recorded by special thermometers called minimum and maximum thermometers. These are usually mounted as shown in figure 146. The upper one is the minimum and the lower the maximum thermometer. In the **maximum** thermometer, the bore is constricted just above the bulb, so that the mercury passes through with some difficulty when the temperature rises and does not run back again when the temperature falls. The **minimum** thermometer is filled with alcohol, and contains within its tube a small black index rod, which is shaped like a double-headed pin. As the temperature falls, the index is drawn down toward the bulb by the surface of the alcohol, and when the temperature rises, the index is left behind.

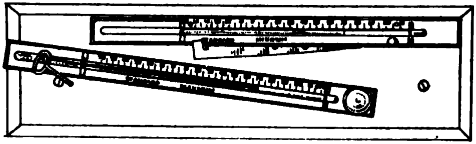


FIG. 146. — Minimum and maximum thermometers.

Another kind of thermometer, which is used by doctors and nurses to detect fever, is the **clinical** thermometer (Fig. 147). This is a maximum thermometer on the Fahrenheit scale, and the range is from 92° F to 110° F, each degree being divided into fifths. The normal temperature of the human body is 98.6° F or 37° C.

FIG. 147. Clinical thermometer.

QUESTIONS AND PROBLEMS

1. Change to Centigrade: 70° F, 150° F, 0° F, −10° F.
2. Change to Fahrenheit: 15° C, 500° C, −26° C, −190° C.

3. What would a rise in temperature of 80° on the Centigrade scale be in Fahrenheit divisions?

4. The temperature of the air on a certain day was 90° F at noon and 45° F late the next night. What was the "drop" in Centigrade degrees?

5. At what temperature do a Centigrade and a Fahrenheit thermometer read the same?

6. How do primitive people start a fire?

7. Why do sparks fly from car wheels when the brakes are quickly applied?

8. Why must a tool be kept wet with cold water when being sharpened on a grindstone?

9. After violent physical exercise one feels very hot. Is the body temperature higher than normal?

10. If one wants the division marks far apart on the stem of a thermometer, what must be the relative size of bulb and stem?

167. Expansion by heat — Solids. When a railroad track is built, a gap is usually left between the ends of the rails, to allow for the expansion of the steel in summer. Iron rims are placed on wheels while hot, because they are then bigger and can be easily slipped on. When they cool, they

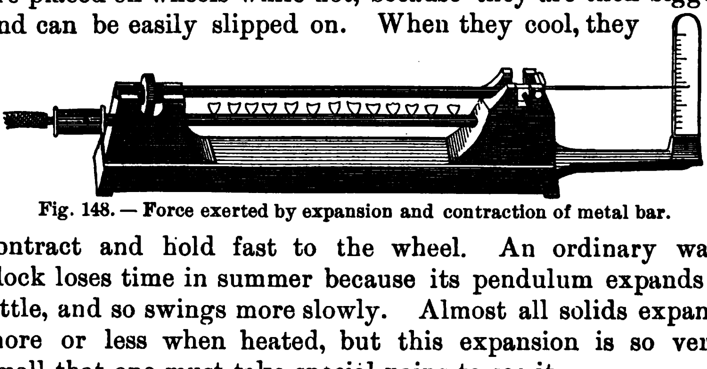


Fig. 148. — Force exerted by expansion and contraction of metal bar.

contract and hold fast to the wheel. An ordinary wall clock loses time in summer because its pendulum expands a little, and so swings more slowly. Almost all solids expand more or less when heated, but this expansion is so very small that one must take special pains to see it.

When solids expand and contract, they may exert enormous forces. We can show in a striking way the power exerted by the expanding and contracting of a metal bar in the following experiment.

First, let us show that a metal bar does expand when heated. In the apparatus shown in figure 148, there is a metal bar which is heated by a

series of little flames below. The expansion, although very slight, is shown by the bent lever at the end, so that as the bar gets hot, the pointer rises. Second, let us show the great force exerted by this process. If we put a steel rod in the slot at right angles to the bar near the lever, when we heat the metal bar, the steel rod suddenly breaks and the pointer is thrown violently up. If we put another steel rod through a hole in the bar, and allow the bar to contract, the steel rod suddenly snaps and the pointer is thrown violently down.

Careful experiments show that different metals expand at different rates. Platinum, for example, expands less and zinc more than other common metals. If we made a platinum meter rod correct at 0° C, it would be 0.9 millimeters too long at 100° C. Similarly a steel meter rod would be 1.3 millimeters too long, and a zinc meter rod would be 2.9 millimeters too long. If two different metal strips, such as iron and brass, are riveted together (Fig. 149), forming a compound bar, the bar when heated will bend or curl, because of the unequal expansion of the metals. Which metal will be on the inner side of the arc? Such compound bars are often used to regulate the temperature of chicken incubators.

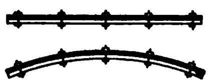


Fig. 149. — Effect of heating a compound bar.

168. Measurement of expansion. In considering how much a given object — such as a steel rail — will expand, it is necessary to know three things about it, namely, its length, and the rise in temperature and the rate of expansion of the particular substance used. For example, if we know that a steel rail is 33 feet long and each foot of it expands 0.000013 feet per degree Centigrade, we can compute how much it will expand from winter to summer, a range of perhaps 50° C. The expansion is equal to the expansion per degree for one foot, multiplied by the length in feet and by the rise in temperature. That is,

$$\begin{aligned}\text{Expansion} &= 0.000013 \times 33 \times 50 \\ &= 0.0214 \text{ feet} = 0.257 \text{ inches.}\end{aligned}$$

We can express this in the form of an equation, thus,

$$e = kl\,(t' - t),$$

where e = expansion,
k = expansion per degree, per unit length,
l = length,
t' = temperature when hot,
t = temperature when cold.

The factor k is called the **coefficient of linear expansion.** It is a very small fraction, and it varies with different substances. It should be remembered that no matter in what unit l is expressed, e will come out in the same unit. Usually k is given per degree Centigrade, but the coefficient for the Fahrenheit scale can be computed by multiplying by $\frac{5}{9}$. Why?

The coefficients per degree Centigrade of some common substances are given in the following table: —

Zinc	0.000029	Steel	0.000013
Lead	0.000029	Cast iron	0.000011
Aluminum	0.000023	Platinum	0.000009
Tin	0.000022	Glass	0.000009
Silver	0.000019	"Invar" (nickel steel)	0.0000009
Brass	0.000018		
Copper	0.000017		

169. Some illustrations. In the construction of a steel bridge allowance has to be made for the expansion of the steel. For example, in the great bridge over the Firth of Forth in Scotland, which is over a mile and a half long, the total expansion amounts to 6 feet. In steam plants, long pipes are provided with sliding or "expansion" joints, unless the bends in the pipe are such as to yield enough for the expansion.

When a lamp chimney is hot, the glass expands. If a

drop of water strikes it, the glass in the immediate vicinity cools rapidly and pulls away from the rest, cracking the chimney.

Quartz is made into crucibles and other objects that are as clear as glass, but have so small a coefficient of expansion (0.0000005) that a red-hot crucible may be suddenly thrust into water without cracking.

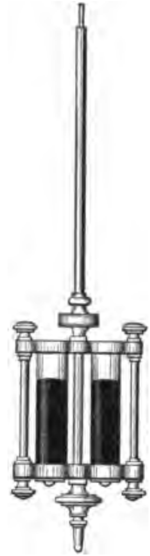


FIG. 150. — Compensated mercury pendulum.

The pendulum rod of a clock is often made of dry wood, which expands very little. It is, however, affected by moisture; so for the most accurate clocks some kind of a compensated metallic pendulum is used. One form of compensated pendulum is that commonly seen in the so-called French clocks. It consists of a glass tube or tubes filled with mercury (Fig. 150), suspended by a steel rod. When properly adjusted, the raising of the center of gravity of the mercury, due to its expansion, is equal to the lowering of the whole reservoir of mercury due to the expansion of the steel rod, so that the effective length of the pendulum remains constant.

In a watch, the balance wheel if uncompensated will run slower in hot weather because the hairspring has less elasticity at a higher temperature, and also because the expansion of the radius of the wheel carries the rim farther from the center, and so slows down its rotation. The rim is therefore made of two strips of metal, brass on the outer edge and steel on the inner, fastened with screws as shown in figure 151. When the temperature rises, the free ends of the rim curl inward, thus bringing part of the rim nearer the axis. This compensates for the expansion of the crossbar and the weakening of the hairspring.

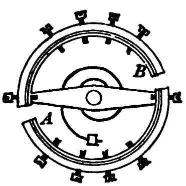


FIG. 151. — Balance wheel of a watch.

PROBLEMS

1. A brass meter bar is correct at 15° C. What will be the error at 20° C?

2. A steel rail 30 feet long is found to expand 0.235 inches when heated from −17° F to 100° F. What is the coefficient of linear expansion on the Fahrenheit scale, and also on the Centigrade scale?

3. The steel cables of a suspension bridge are 2000 feet long. How much do they change in length between the temperatures −20° F and 97° F?

4. A steel shaft is heated to 65° C while being shaped in a lathe, and its diameter at that temperature is made just 5 centimeters. What will its diameter be at room temperature (15° C)?

5. A steel wire, 150 centimeters long at 15° C, becomes 151.3 centimeters long when an electric current is sent through it. How hot does it get?

170. Cubical expansion of solids. A metal bar when heated expands, not only in length, but also in breadth and thickness; in short, its volume increases. This expansion in volume is called **cubical** expansion. Suppose we have a cube 1 centimeter on an edge at 0° C and raise its temperature to 1° C; each edge of the cube will become $(1+k)$ centimeters, k being the coefficient of linear expansion. The original volume, 1 cubic centimeter, will become $(1+k)^3$ cubic centimeters. Now $(1+k)^3$ equals $1+3k+3k^2+k^3$; but since k is a very small fraction, the value of $3k^2$ and k^3 will be so small that they may be neglected without appreciable error. The volume of the cube is, then, $1+3k$; hence the volume expansion per cubic centimeter per degree is $3k$ cubic centimeters and *the coefficient of cubical expansion is three times the coefficient of linear expansion.*

For example, the coefficient of linear expansion of glass is 0.000009, and the coefficient of cubical expansion is 3 times 0.000009 or 0.000027. A flask which held just a liter at 0° C would hold 1002.7 cubic centimeters at 100° C.

171. Expansion of liquids. Let us fill a small round-bottomed flask with water colored with ink and insert a stopper with a glass tube

and paper scale (Fig. 152). Then let us put the flask into a jar of ice water and mark on the scale the position of the liquid in the tube. If we then put the flask into a basin of boiling water, we shall note at first a sudden drop of the liquid in the tube (why?) and then a rapid rise. Evidently the liquid expands more than the glass.

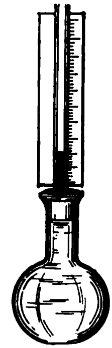


Fig. 152. Expansion of a liquid.

In general it is found that liquids expand much more than solids. For example, when a liter of water is heated from 0° to 100° C, it increases in volume about 40 cubic centimeters, whereas a block of steel of the same volume would expand only 3.9 cubic centimeters. Alcohol, oils, and especially kerosene expand even more than water.

Liquids, like solids, expand with almost irresistible force when heated, and exert enormous pressures if expansion is prevented by their surroundings.

In the case of liquids and gases, cubical expansion rather than linear is what is always measured. Since, however, the vessel which contains the liquid expands as well as the liquid, we observe only the *apparent* expansion. In a mercury thermometer the apparent expansion is only about $\frac{5}{6}$ of the real expansion of the mercury. The coefficient of cubical expansion of alcohol is 0.00104, of mercury 0.000181, and of water from 0.000053 to 0.00059 according to the temperature.

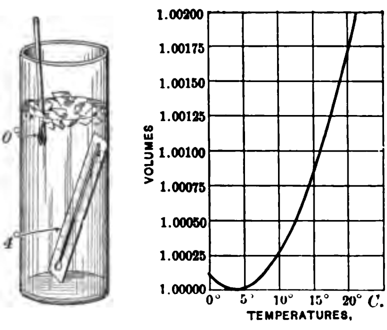


Fig. 153. — Maximum density of water.

172. Abnormal behavior of water. We have just seen that solids, liquids, and gases expand as a rule when heated; water does the same except near its freezing point.

If we fill a tall glass jar nearly full of cracked ice (Fig. 153) and let it stand for a while, the temperature of the water near the top comes to 0° C and remains so, while the temperature at the bottom will be about 4° C. Since the heaviest liquid stays at the bottom, this means that water at 4° C is denser than water at 0°.

Very precise measurements show that *water is most dense at 4° C. When water at 4° C is either warmed or cooled, it expands and becomes lighter, as shown by the curve in figure 153.* This fact has many important consequences. For example, if it were not for this, the water in lakes would freeze in winter, not merely at the surface, but solidly from top to bottom, thus destroying all aquatic life.

FIG. 154. — Expansion of gas.

173. Expansion of gases. We may easily show the great expansion of a gas when heated, with the apparatus shown in figure 154. Even the heat of the hand on the flask causes bubbles of air to be expelled from the tube and to rise through the water. If the heat of a flame is applied to the flask, the bubbles rise rapidly. If after a time the flame is removed and the flask allowed to cool, water rises into the flask to take the place of the escaped air. From the volume of water thus drawn up into the flask, it is evident that a considerable fraction of the air was expelled during the expansion.

The expansion of gases, such as air, illuminating gas, or acetylene, is remarkable for two reasons: first, because it is so large — being about nine times as much as for water, and second, because it is nearly the same for all gases.

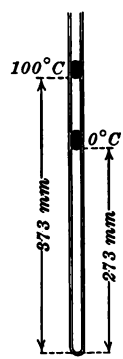


FIG. 155. — Expansion of a gas under constant pressure.

The coefficient of expansion of a gas can be measured as follows. Suppose we have a tube of uniform bore (Fig. 155), which is closed at one end and has a little pellet of mercury to separate the inclosed gas

from the atmosphere. (Dry air is a good gas to experiment with.) If we put the tube in a freezing mixture at 0° C, the gas in the tube will contract, and we can measure the length, which we will suppose is 273 millimeters. If we put the tube in steam at 100° C, the gas will expand, and we can measure the length again. We shall find that it is about 373 millimeters. From this it is evident that the gas has expanded 1 millimeter for each degree rise in temperature (the expansion of the glass can be neglected). That is, it has expanded $\frac{1}{273}$ or 0.00366 of its volume at 0° C for each degree rise in temperature.

Gay-Lussac (1778–1850) was one of the first to study the expansion of gases under constant pressure. He found that *different gases have nearly the same coefficients of expansion, namely* $\frac{1}{273}$ *or 0.00366.*

174. Pressure coefficient of gases. Since the volume of a gas increases as the temperature rises, it is reasonable to expect that if a certain quantity of gas were heated and yet confined in the same space, the pressure would increase. The following experiment shows that this is true.

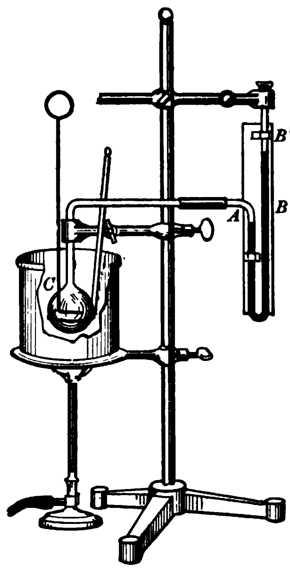


FIG. 156. — Pressure of gas, heated at fixed volume, increases.

Let us start with a gas like dry air, confined in a bulb *C*, which is connected with an open manometer *AB*, as shown in figure 156. At first we will surround the bulb by melting ice, so that the gas is at 0° C, and have the mercury at the same level in each arm of the manometer, so that the gas is at atmospheric pressure. Then we will surround the bulb with boiling water at 100° C, and keep the gas from expanding by pouring mercury into the manometer arm *B*, thus increasing the pressure. This increase of pressure is measured by the difference in levels *B'* and *A*. From this we may calculate the increase per degree rise in temperature, and finally what fraction it is of the pressure at 0° C. The result is called the **pressure coefficient** of the gas.

Very careful experiments of this sort were first carried out by a Frenchman, Regnault (1810–1878), who found that *the pressure of a gas kept at constant volume increases for each degree very nearly $\frac{1}{273}$ or 0.00366 of the pressure at 0°, no matter what the gas is.* It will be noticed that this is the same fraction which we found for the increase of *volume.*

To sum up —

I. *Different gases have nearly the same coefficients of expansion;*

II. *Different gases have nearly the same pressure coefficients;*

III. *The pressure coefficient of any gas is numerically about the same as its coefficient of expansion; each is about $\frac{1}{273}$ or 0.00366.*

175. Gas thermometers. It is evident that by measuring the increase of volume of a gas under constant pressure or the increase of pressure of a gas kept at constant volume, we have a means of measuring temperature changes. *Such a thermometer, filled with hydrogen, is used as the world's standard thermometer at the International Bureau of Weights and Measures near Paris.* Since the hydrogen thermometer has been chosen as the standard, it is important to know just how closely a good mercury thermometer agrees with it. Of course they agree exactly at the two fixed points 0° and 100° C, and a careful comparison shows that between 0° and 100° the difference is not over 0.12° at any point.

176. Absolute temperature scale. In the experiment described in section 173, we started with an air column 273 millimeters in length at 0° C; if we had cooled the gas from 0° to −1° C, the length *AB* would have been shortened a millimeter, and if we had cooled it to −10° C, the length of the air column would have become 263 millimeters. If, then, the air column continued to contract at the same rate if cooled indefinitely, the volume of the air at −273° C would be zero.

As a matter of fact, we can never get a gas to so low a temperature as $-273°$ C, for every known gas, before that temperature is reached, becomes a liquid. This temperature **$-273°$ C** is, however, one of unusual interest in the study of gases. It is called the **absolute zero**, and temperatures measured from this point as zero are called **absolute temperatures**. Absolute temperatures may be designated by the letter A. Thus, 0° C is 273° A, 50° C is 323° A, and 100° C is 373° A. To change any temperature from the Centigrade to the absolute scale, we have merely to add 273 degrees (Fig. 157).

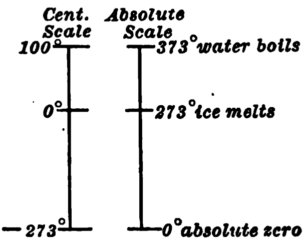


FIG. 157. — Absolute and Centigrade scales.

From the above discussion of absolute temperature it will be seen that the volume of any gas is doubled when its temperature is raised from 273° A (0° C) to $2 \times 273°$, or 546° A (273° C). *In general, the volume of a gas is very nearly proportional to its absolute temperature when the pressure is kept constant.*

Now since, by section 174, the coefficient of expansion of a gas at constant pressure is the same as the pressure coefficient at constant volume, *when the volume is kept constant, the pressure of a gas is proportional to the absolute temperature.*

177. Gas formula. The relation between the volume and the temperature of a gas can be very concisely expressed algebraically, thus,

$$\frac{V}{V'} = \frac{T}{T'}, \tag{I}$$

where V and V' represent the volumes of a certain gas at the same pressure, but at different absolute temperatures T to T'. If t is the temperature on the Centigrade scale when the volume is V, then $T = 273 + t$; similarly $T' = 273 + t'$.

The relation between the volume and pressure of a gas at

constant temperature may be concisely expressed by Boyle's law (see section 87),

$$PV = P'V', \tag{II}$$

where V is the volume of a given quantity of gas under a pressure P, and V' is the volume of the same gas under a pressure P', the temperature in the two cases being the same.

The relation of the volume to both pressure and temperature can be expressed by the equation,

$$\frac{PV}{T} = \frac{P'V'}{T'}; \tag{III}$$

for it is readily seen that this equation reduces to equation (II), if $T = T'$, and that if $P = P'$, the equation becomes $V/T = V'/T'$, which is another form of equation (I). Equation (III) is called the **gas formula.**

A problem will make clear the use of the gas formula. Suppose we wish to find the volume of a certain quantity of gas under standard conditions, that is, at 0° C, and 760 millimeters pressure, when it is known to occupy 120 cubic centimeters at 15° C and under a pressure of 740 millimeters. Substituting in equation (III), we have

$$\frac{120 \times 740}{273 + 15} = \frac{V' \times 760}{273 + 0},$$

whence

$$V' = 111 \text{ cubic centimeters.}$$

Problems

1. At what temperature on the Centigrade scale will a liter of air at 0° expand to occupy 2 liters, the pressure being held constant?
2. A certain quantity of gas occupies 350 cubic centimeters at 27° C. What will be its volume at 0° C, the pressure being held constant?
3. A steel tank full of air at 15° C under atmospheric pressure was sealed and thrust into a furnace, where it was heated to 1000° C. How many atmospheres of pressure did the air then exert? Neglect the thermal expansion of the steel.
4. A liter of air at 0° C and atmospheric pressure weighs 1.293 grams. What is the density of air at 100° C and atmospheric pressure?

5. A student in a chemical laboratory generates 50 liters of hydrogen at 10° C, and at a pressure of 700 millimeters. Find the volume of the gas under standard conditions; that is, at 0° C and at 760 millimeters.

6. At the beginning of the so-called "compression stroke" in an automobile engine, its cylinder contains 42 cubic inches of gas and air at atmospheric pressure, and at a temperature of 40° C. At the end of the compression the volume is 6 cubic inches and the pressure is 15 atmospheres. What is the temperature?

178. Low temperatures. The investigations of Lord Kelvin (1824–1907) and of other scientific men all point to the conclusion that the temperature − 273° C is really an **absolute zero** in the same sense that it is **the lowest possible temperature in the universe.** Although no one has as yet succeeded in cooling a body to absolute zero, temperatures within a very few degrees of this point have been attained by the evaporation of liquefied gases. With liquid air, temperatures as low as − 200° C may be obtained, and with liquid hydrogen − 258° C. In 1908 Professor Onnes, at the University of Leyden in Holland, found that the boiling point of liquid helium is − 268.6° C, or only about 4.5° above the absolute zero, and he has since cooled liquid helium to within 2° of the absolute zero. At these low temperatures rubber and steel become as brittle as glass, and metals become much better conductors of electricity than at ordinary temperatures.

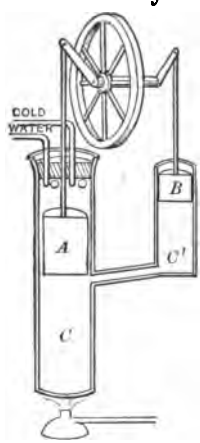


FIG. 158. — Diagram of hot-air engine.

179. Hot-air engine. An interesting application of the expansion of gases is the hot-air engine. Its operation can be best understood by studying figure 158. A loosely fitting plunger *A* moves up and down and thus shifts the air back and forth in the cylinder *C*, which is heated at the bottom and kept cool at the top. The working cylinder *C′* has a nicely fitting piston *B*.

When the plunger A moves down, the hot air below is transferred to the top, where it is cooled. This makes it contract. The piston B is then forced down by the external pressure of the atmosphere. As soon as the piston B is near the bottom of its stroke, the plunger A is raised, causing the air to flow back under A, where it is heated by the fire. This makes it expand and forces the piston B up again, and so the cycle is repeated.

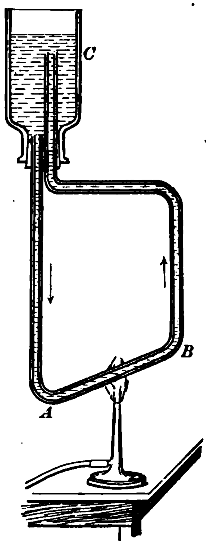


FIG. 159. — Convection current of water.

These engines are commonly used for pumping water on a small scale at isolated places, for they do not require expert attendants, and they use any kind of fuel. In general they cannot compete with gas engines on account of their bulk and the rapid wearing out of the heating surfaces.

180. Convection currents. All systems of heating and ventilation depend upon what are called convection currents, which in turn depend upon the expansion of liquids and gases. To make these clear, let us try two simple experiments.

We cut off the bottom of a bottle and bend a glass tube (Fig. 159) so that the ends can be slipped through a stopper which fits the neck of the bottle. If we invert the bottle and fill it with water containing a little sawdust, we can see a circulation of the water when a flame is waved back and forth from A to B. We note that the direction is from A to B. Why?

A box (Fig. 160) has a glass front, and two holes in the top, which are covered with glass chimneys. If we put a candle under one chimney, convection currents

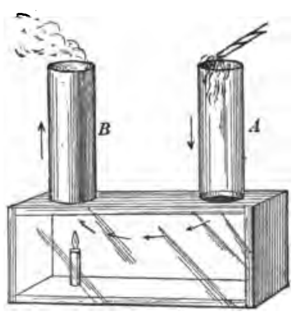


FIG. 160. — Convection current of air.

of air go down the cool chimney and up the warm one. A bit of lighted touch paper held near the top of the cool chimney makes the convection currents more evident.

The draft in a lamp, stove, fireplace, or power-house chimney is a convection current.

The explanation of the movement of convection currents is that any gas or liquid expands when heated, so that a given quantity of fluid increases in volume and consequently decreases in density. In a convection current, the lighter fluid is pushed up by the heavier surrounding fluid, just as a block of wood under water is pushed up by the surrounding water.

Transmission of Heat

181. Heat transfer by convection. Since the up-going part of a convection current is warmer than the returning part, there is a transfer of heat from the flame or other source of heat at the bottom, to the cooler parts of the circuit at the top. This process of transporting heat by carrying hot bodies or hot portions of a fluid from one place to another is called **convection.** It is the basis of almost all systems for heating buildings.

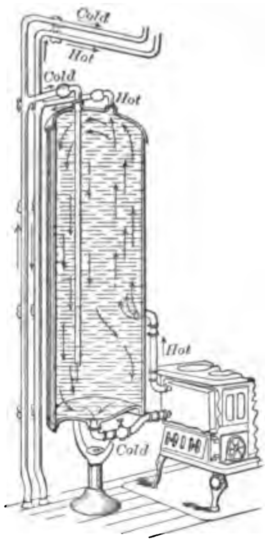


Fig. 161.— Hot-water heater.

182. Hot-water heating. The arrangement for **heating water in the kitchen range** for general use in laundry and bathroom is shown in figure 161. The cold water enters the tank through a pipe which reaches nearly to the bottom. From the bottom of the tank the water is led to a heating coil along the side of the fire box in the range. When the water becomes hot, it is pushed up and goes back into the

tank at a point nearer the top. Thus a circulation is set up which continues until practically all the water in the tank has passed through the stove and the whole tankful is hot.

The **hot-water system of heating houses** depends on this same principle of convection. Water is heated nearly to the boiling point in a furnace in the basement. The hot water is led from the top of the furnace through pipes to iron radiators in the various rooms of the building. On account of the large exposed surface in each radiator, heat is rapidly given out by the hot water to the surrounding air. The cooled water is then carried from the radiators through return pipes to the base of the furnace. To prevent radiation from the pipes, a thick non-conducting coating of asbestos is often provided.

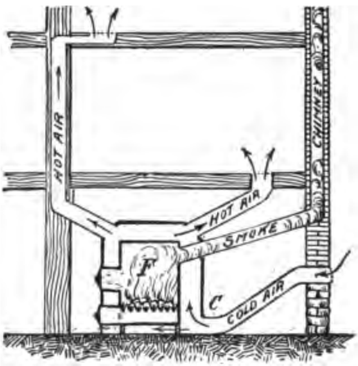


FIG. 162. — The hot-air furnace.

163. Hot-air system of heating and ventilating. The **hot-air furnace** in the basement (Fig. 162) is simply a big stove, surrounded by a shell or jacket of galvanized sheet iron. The air between the stove and outer shell is heated, and is then pushed up into the flues by the heavier cold air which comes in from out of doors through the cold-air inlet flue. The smoke, of course, goes up the chimney. The warm air which enters the rooms finds an outlet around the doors and windows.

In the hot-water system of heating there is no provision whatever for changing the air in the room; that is, for **ventilation.** In the hot-air system, a small quantity of fresh air is continually flowing into the rooms. This is enough for a private house. But in schools, churches, and other public buildings, large quantities of clean, fresh, warm air have to

be continually supplied by other means. For the proper ventilation of a room it is estimated that each person in it requires about 50 cubic feet of fresh air every minute. In large modern school buildings the air is drawn in from out of doors by powerful fans, filtered through cloth, warmed by passing around steam pipes, and then distributed in ducts throughout the building. The vitiated air in each room is forced out through ducts near the floor. This **indirect system of heating**, while expensive, furnishes excellent ventilation.

184. Conduction in solids. Besides transporting heat from one place to another by carrying hot bodies about, or making hot fluids flow through pipes, we can **transmit** heat from one place to another, without moving any material thing, by either of two methods called **conduction** and **radiation.**

Every one knows that the handle of a silver spoon gets hot when its bowl is in a cup of hot tea or coffee. If one end of an iron poker is put in the fire, the other end gets uncomfortably hot and must be provided with a wooden handle. Yet if a wooden rod is plunged into a fire, it is hard to feel any warmth at the other end. So we conclude that silver and iron conduct heat better than wood. In general, metals are good conductors of heat.

There are some substances, such as stone, glass, wood, wool, fur, and ashes, which are poor conductors of heat and are therefore called heat **insulators.** The metals, such as silver, copper, brass, iron, lead, etc., are good **conductors** as compared with the non-metals. Careful study shows that even the metals vary in their power to conduct heat, that is, in conductivity.

This can be shown by the following experiment.

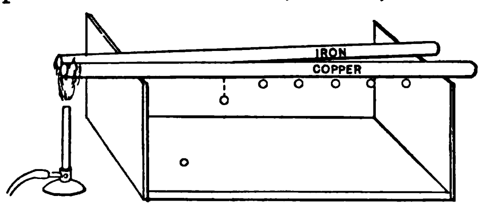


FIG. 163. — Relative conductivity of copper and iron.

Let us fasten with sealing wax a number of steel balls at regular

intervals on the under side of two rods, one of copper and the other of iron. If we heat one end of each rod in a flame (Fig. 163), the balls on the copper rod soon begin to drop off, beginning near the flame. Later the balls on the iron rod begin to drop off. Often half the balls will have dropped from the copper rod before the first one drops from the iron rod.

185. Conduction in liquids and gases. Liquids and gases are much poorer conductors than metals. This can be shown by the following experiments.

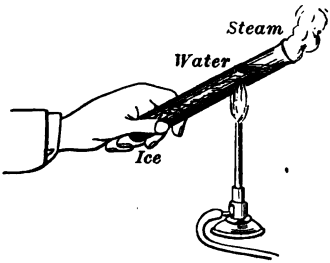


FIG. 164. — Water a non-conductor.

Let us take a test tube full of water and place in it a few pieces of ice which are held in the bottom by a wire, as shown in figure 164. Then we may boil the water at the top of the tube for some time without melting the ice in the bottom.

Another more striking experiment to show the poor conductivity of water is shown in figure 165. The bulb of the air thermometer is placed only half an inch below the surface of the water in the funnel. When a spoonful of ether is poured on the surface of the water and lighted, the liquid in the tube of the air thermometer will remain practically stationary, in spite of the fact that the air thermometer is very sensitive to changes in temperature.

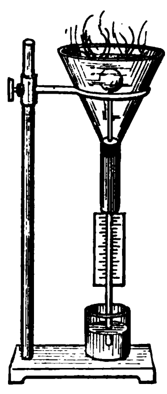


FIG. 165. — Burning ether on the water does not affect the air thermometer.

Experiments to measure conductivity show that iron conducts 100 times as well as water, and that water conducts 25 times as well as air. In general, it may be said that liquids and gases are very poor conductors of heat.

It is an interesting fact that substances which are good conductors of heat are good conductors of electricity as well.

186. Applications. These differences in conductivity explain why teapots have wooden or insulated handles; why steam

pipes are covered with wool, magnesia, or asbestos;. why double windows are used in cold climates; why a vacuum bottle (Fig. 166) keeps things hot or cold; and why we wear woolen clothing in winter. Woolen clothing of loose texture, furs and feathers, or eiderdown quilts are effective as heat insulators because so much air is inclosed in their pores.

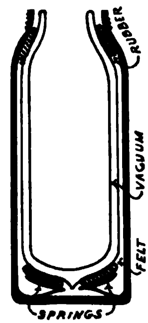


FIG. 166.—Section of vacuum bottle: a vacuum is a very poor conductor of heat.

Differences in conductivity also account for many of our curious sensations of heat and cold. Thus in a cool room some things feel much colder than others. Metallic objects, which are good conductors, take heat rapidly from the hand, and so give the sensation of cold. While other objects, such as wood and paper, do not carry off the heat of the hand and so do not feel cold. Similarly a piece of metal lying in the hot sun feels much warmer than a piece of wood beside it.

187. Radiation. If an iron ball is heated and hung up in the room, the heat can be felt when the hand is held *under* the ball. This cannot be due to convection, because the hot-air currents would rise from the ball. It cannot be due to conduction because gases are very poor conductors. Similarly a lighted electric light bulb feels hot if the hand is held near it, but when the light is turned off, the sensation stops very quickly. The glass of the bulb is a very poor conductor and there is practically no air left inside the bulb, so that the sensation of heat can be due neither to convection nor to conduction. Furthermore, an enormous quantity of heat comes to us from the sun. Yet men who make ascents in balloons and aëroplanes find that the air becomes less and less dense, so that it seems reasonable to suppose that the earth's atmosphere forms a coating only a few miles thick

and that the space beyond is absolutely empty. So the sun's heat cannot come by convection or conduction.

Scientists, to explain these phenomena, have imagined a weightless, elastic fluid called **the ether** which fills all space and transmits heat and light by a process called **radiation.** When a body not in contact with conducting bodies cools, it is said to radiate heat, or to cool by radiation. If one places a screen, such as a book, between a lighted lamp and his face, he no longer feels the heat. So we think that heat rays, like light rays, travel in straight lines. Experiments also show that heat rays, like light rays, can be reflected by a mirror, or brought to a focus by a burning glass.

Some substances, such as glass and air, let the sun's heat rays pass through almost unimpeded and are warmed but little by this radiant heat; that is, they are "**transparent-to-heat.**" Other substances, such as water, do not let heat pass through and are warmed by any radiant heat rays that strike them; they are "**opaque-to-heat.**"

A mirror, or any highly polished surface, is a good heat reflector, and yet itself remains cold. Fresh snow melts slowly in the sun's rays, but snow covered with soot or black dirt melts rapidly. *In general, reflecting or white objects do not easily absorb radiant heat, while rough or black objects absorb heat readily.*

It has also been found that reflecting and bright-colored objects, when hot, cool by radiation more slowly than rough and dark objects. For example, a brightly polished silver cup radiates heat twenty times more slowly than a sooty black cup. *In general, good absorbers are good radiators, and poor absorbers are poor radiators.*

188. Theory as to what heat is. There are many reasons for thinking that heat is a rapid vibratory motion of the molecules of substances or of the ether which fills the spaces between the molecules. We imagine that the molecules in a hot flatiron are vibrating more rapidly than when it was

cold, and that this molecular vibration extends to the surrounding ether and so is sent out in straight lines in all directions as radiant heat.

At a temperature of about 550° C iron becomes "red hot," and at 1300° C it gets "white hot." We imagine that the iron, before it begins to glow, is sending out dark heat rays, but that, when red hot or white hot, it is sending out visible heat rays, that is, light rays. We think that these heat rays and light rays differ only in the rapidity of the vibratory motion, and in their effect on man's organs of sense. If the vibrations are under 400 trillion per second, we recognize them as *heat;* but if the vibrations are between 400 and 800 trillion per second, the nerves of the eye recognize them as *light*. Heat and light are both forms of radiant energy. This radiant energy travels at the enormous speed of 187,000 miles per second, which means that radiant energy could circle the earth seven times in one second.

On this theory, the expansion of bodies when heated is due to the more violent vibration of their molecules, which require more room to move about in. At a certain temperature this motion becomes so violent that the molecules break away from their former position and the body changes its state; that is, it melts or boils.

SUMMARY OF PRINCIPLES IN CHAPTER X

100 Centigrade degrees = 180 Fahrenheit degrees.

$$\text{Temp. Fahr.} = \left(\tfrac{9}{5}\ \text{Temp. Cent.}\right) + 32.$$

$$\text{Temp. Cent.} = \tfrac{5}{9}\ (\text{Temp. Fahr.} - 32).$$

$$\text{Coefficient of linear expansion} = \text{expansion per degree for unit length}$$
$$= \frac{\text{total expansion}}{\text{total length} \times \text{rise in temperature}}.$$

$$\text{Total expansion} = \text{coefficient} \times \text{length} \times \text{rise in temperature.}$$

o

Coefficient of volume expansion

$$= \text{expansion per degree for unit volume}$$

$$= \frac{\text{total expansion}}{\text{total volume} \times \text{rise in temperature}}.$$

Total expansion = coefficient × volume × rise in temperature.

For solids, volume coefficient = 3 × linear coefficient.

$$\textbf{Pressure coefficient of gases} = \frac{\text{total pressure rise}}{\text{pressure} \times \text{rise in temperature}}.$$

Total pressure rise = coefficient × pressure × rise in temperature.

$$\left.\begin{array}{l}\textbf{Volume coefficient of all gases nearly the same.}\\ \textbf{Pressure coefficient of all gases nearly the same.}\\ \textbf{Volume and pressure coefficients nearly equal.}\end{array}\right\} \textbf{Value } \frac{1}{273}.$$

Gas law: $\dfrac{PV}{T} = \dfrac{P'V'}{T'}$.

QUESTIONS

1 Is friction ever a source of useful heat?

2. Are the sun's rays ever used practically as a direct source of heat for engines?

3. Why does spring water seem warm in winter and cool in summer?

4. Why does the water seem much colder before a bath than afterwards?

5. Why can a platinum wire be sealed or melted into glass while a copper wire cannot?

6. Why do glass bottles crack when placed on a hot stove?

7. Why do apples and pieces of green wood swell when heated?

8. Why is there a cold indraft of air at the bottom of an open window?

9. Is there any other reason than convenience for putting furnaces in cellars rather than in attics?

10. How is the water which is standing in the hot-water pipes in a house kept hot?

11. Does a hot body cool more rapidly if placed on metal than if placed on wood? Why?

12. Why does a glowing coal die out quickly on a metal shovel, and yet glow for a long time in ashes?

13. How does a fireless cooker work?

14. Look up Davy's lamp for miners in an encyclopedia. What is its advantage? Why is it that a flame will not strike through the fine-net wire gauze?

15. Why are the walls of ice houses often packed with sawdust?

16. Why should an air space be left in building the walls of brick and cement houses?

17. Does woolen clothing supply any heat to maintain the body's temperature?

18. Why do people prefer to wear white clothes in summer and in hot countries?

19. Why should the surface of a teakettle be brightly polished and the bottom blackened?

20. Is it advisable to put any sort of aluminum or gold paint on a radiator that is to heat a room?

21. Describe carefully the "dampers" of some stove or furnace you have seen, and explain how they accomplish the desired results.

CHAPTER XI

WATER, ICE, AND STEAM

Measurement of heat — B. t. u. and calorie — specific heat — freezing point — change of volume in freezing — latent heat, ice to water — boiling point under various pressures — distillation — latent heat, water to steam — humidity — fog, rain, and snow — artificial ice.

189. How we measure heat. If a man buys a ton of coal, what does he get for his money? One answer would be, about 2000 pounds of material, of which, perhaps, 40 pounds is water, 320 pounds is ash, and the rest mostly carbon and hydrogen. What the man is really interested in, however, is not the sort of material, but the amount of heat he has bought. Since heat is not a substance, but a form of energy, we cannot measure it directly in pounds or quarts, but must measure it by the effect it can produce. For example, if one pound of hard coal could be completely burned, and if all the heat generated in this process could be used to heat water, it would be found that about 7 tons of water could be raised 1° F in temperature. Engineers reckon the heat value of fuel in units such that each represents the *heat required to raise one pound of water one degree Fahrenheit.* This heat unit is called the "**British thermal unit,**" and is written B. t. u. For example, the heat value of a pound of coal varies from 11,000 to 16,000 B. t. u.; a pound of petroleum gives about 25,000 B. t. u., a pound of gasolene about 19,000 B. t. u., and a pound of dry wood about 5000 B. t. u.

The heat unit employed in Europe, and in all physical and chemical laboratories, is a metric unit called the **calorie.** *The calorie is the heat required to raise the temperature of a gram of water one degree Centigrade.*

190. Heat absorbed by different substances. It is well known that a kettle of water on a stove gets warm much less quickly than a flatiron of the same *weight*. For example, the heat required to warm a kilogram of water 1 degree will warm the same *weight* of copper 10 degrees, of silver or tin 20 degrees, and of lead or mercury 30 degrees. In fact experiments show that water requires more heat per unit weight per degree rise of temperature than any other common substance.

Since one calorie is required to raise the temperature of one gram of water one degree, only one tenth of a calorie would be needed to raise the temperature of one gram of copper a degree, one twentieth of a calorie to raise a gram of silver one degree, and one thirtieth of a calorie to raise a gram of lead one degree. *The number of calories required to raise the temperature of a gram of a substance one degree Centigrade is called its* **specific heat.** Thus the specific heat of water is 1, of copper about 0.1, etc.

The following experiment of Tyndall's illustrates how much substances differ in their specific heats.

We may heat a number of balls of the same weight but of different metals, such as iron, zinc, copper, lead, and tin, to about 150° C in oil. Then if we place them all at the same time on a thin cake of paraffin wax which is held on a ring, as shown in figure 167, they will melt the wax and sink into it, but at different rates. The iron works its way most vigorously into the wax, and even through the cake. The zinc and copper balls come next, while the lead ball makes but little headway. The metal with the largest specific heat, iron, gives out the largest amount of heat in cooling and so melts the most paraffin.

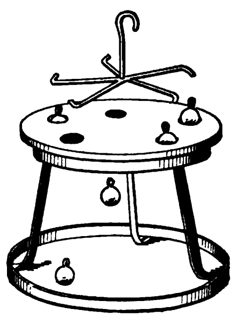


Fig. 167. — Metals differ in specific heat.

191 How specific heat is determined. When a hot substance, such as hot mercury, is poured into cold water, the

water and mercury soon come to the same temperature. The heat given up by the cooling mercury is used in warming the water. If no heat is lost in the process, *the heat units given out by the hot body are equal to the heat units gained by the cold body.*

This **method of mixtures** is accurate only when no heat is lost during the transfer. This is rather difficult to manage in practice. Nevertheless, this method is the one generally used in laboratories to determine the specific heat of substances.

For example, suppose that 300 grams of mercury are heated to 100° C and then quickly poured into 100 grams of water at 10° C, and that, after stirring, the temperature of the water and mercury is 18.2° C.

If we let x be the specific heat of the mercury, the mercury gives out $300(100-18.2)x$ calories. Since the specific heat of water is 1, the water absorbs $100(18.2-10)1$ calories. Therefore we may make the equation

$$300(100-18.2)x = 100(18.2-10)1,$$
$$\text{whence } x = 0.033 \text{ calories.}$$

By very careful experiments of this sort the specific heats of some of the common substances have been found to be as follows: —

TABLE OF SPECIFIC HEATS

Water	1.00	Sand	0.19
Pine wood	0.65	Iron	0.12
Alcohol	0.60	Copper	0.094
Ice	0.50	Zinc	0.093
Aluminum	0.22	Mercury	0.033

It is remarkable that of all ordinary substances water has the greatest specific heat. Thus it takes about four times as much heat to raise a pound of water one degree as to raise a pound of solid earth one degree, and so the ocean acts as a great moderator of temperatures. In summer the water absorbs a vast amount of heat which it slowly gives up in winter to the land and air. This explains why the temperature on some ocean islands does not vary more than 10° F during the whole year.

PROBLEMS

1. How many calories of heat are needed to raise the temperature of 10 grams of water 5° C?

2. How many calories are required to heat 15 grams of iron 20° C?

3. Compute the calories given out by a kilogram of copper in cooling from 110° C to 15° C.

4. How many B. t. u. are necessary to heat a 2-pound flatiron from 70° F to 350° F?

5. If the heat value of coal is 14,000 B. t. u. per pound, how many tons of water can be heated from 32° to 212° F by the combustion of one ton of coal in a boiler whose efficiency is 75%?

6. If 400 grams of water at 100° C are mixed with 100 grams of water at 20° C, what will be the temperature of the mixture?

7. If 500 grams of copper at 100° C, when plunged into 300 grams of water at 10° C, raise the temperature to 22° C, what is the specific heat of copper?

8. A piece of iron weighing 150 grams is warmed 1° C. How many grams of water could be warmed 1° by the same amount of heat? (The answer is called the **water equivalent** of the piece of iron.)

9. A 50-pound iron ball is to be cooled from 1000° F to 80° F, by putting it in a tank of water at 32° F. How many pounds of water must there be in the tank?

10. A platinum ball weighing 100 grams is heated in a furnace for some time, and then dropped into 400 grams of water at 0° C, which is raised to 10° C. How hot was the furnace? (Sp. heat = 0.04.)

192. Melting and freezing. If one brings in from out of doors on a cold winter day a pailful of snow or ice and sets it on a stove, he finds that its temperature is at first below 0° C and slowly rises to that point. It then remains stationary, or nearly so, until all the snow is melted. Then the temperature of the water gradually rises. This stationary temperature, where the ice (snow) changed to water, is called the **melting point** of ice, and is 0° C or 32° F.

We may also determine the freezing point of water by making a freezing mixture of cracked ice and salt and placing in it a test tube containing some pure water. The temperature of the water will be observed to fall slowly until the water begins to freeze. Then the temperature remains con-

stant until all the water is frozen. This stationary temperature at which water changes into ice is called the **freezing point** of water, and is 0° C or 32° F.

Substances which are crystalline, such as ice and many metals, change into liquids at a definite temperature, and the melting point of such a substance is the same as its freezing point.

TABLE OF MELTING OR FREEZING POINTS

Platinum	above 1700° C	Tin	232° C
Steel	1300 to 1400° C	Sulphur	115° C
Glass	1000 to 1400° C	Naphthalene (moth balls)	80° C
Copper	1083° C	Paraffin	about 54° C
Gold	1062° C	Ice	0° C
Silver	960° C	Mercury	– 39° C
Lead	327° C	Alcohol	about – 112° C

Non-crystalline substances, such as iron, glass, and paraffin, pass through a soft, pasty stage as the melting point is approached. In the case of some substances, such as the fats, the melting point is not the same as the freezing point. Thus butter will melt between 28° and 33° C and yet solidifies between 20° and 23° C.

There are several alloys of metals which melt at a much lower temperature than any of the metals of which they are made. "Wood's metal" (2 tin + 4 lead + 7 bismuth + 1 cadmium by weight) melts at 70° C, although the lowest melting point of any of its constituents is that of tin (232° C). Wood's metal will melt even in hot water. Such alloys are used to seal tin cans and automatic fire sprinklers. Other similar alloys are used for fusible plugs for boilers.

193. Expansion in freezing. When a liquid freezes, we would naturally expect it to contract, because it would seem that the molecules would be more closely knit together in the solid than in the liquid state. This is generally true. But when we recall that ice floats and pitchers of water are often cracked by freezing, we see that water expands on freezing.

In fact a cubic foot of water becomes 1.09 cubic feet of ice. Cast iron is another substance that expands a little in solidifying, and it is therefore adapted to making castings, for in this way every detail of the mold is sharply reproduced. In making good type we must have a metal which expands a little on solidifying, and so an alloy of lead, antimony, and copper, which has this property, is used.

That the expansive force of water in freezing is enormous can be seen from the following experiment.

Let us fill a cast-iron bomb with water, close the hole with a screw plug (Fig. 168), and put the bomb in a pail of ice and salt. When the water in the bomb freezes, the pressure inside increases more and more, and the bomb eventually explodes.

FIG. 168. — Expansive force exerted by freezing water.

This shows why water pipes burst on nights cold enough to freeze the water in them. A similar process is active every winter in breaking the rocks of mountains to pieces. Water percolates into the crevices, freezes, and expands.

194. Effect of pressure on melting ice. If we suspend a weight of 40 or 50 pounds by a wire loop over a block of ice (Fig. 169), the wire will cut slowly through the ice. The pressure causes the ice to melt under the wire; but the water flowing around the wire freezes again above it, and leaves the block as solid as before.

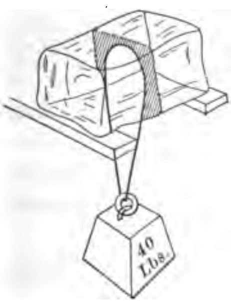


FIG. 169. — Wire cutting through a block of ice.

This experiment shows that pressure causes ice to melt by lowering the freezing point. This might be expected, for pressure on any body tends to prevent its expansion, and since water does expand on freezing, pressure will tend to prevent freezing; that is, it lowers the freezing point. It requires, however, a pressure of almost a ton (1850 pounds)

per square inch to lower the freezing point one degree Centigrade.

In skating, the pressure of the edge of the skate blade melts the ice and so forms a film of water which is very slippery. This also explains how snowballs can be made by pressing the snow between the hands. The pressure at the points of contact between the flakes of snow melts them and then the film of water that is formed freezes again when the pressure is released. The flow of glaciers of solid ice around corners is explained in the same way.

195. Latent heat: ice to water. If a dish of ice and water at 0° C is kept in a room where everything else is at 0°, the ice will not melt and the water will not freeze. But if the dish is surrounded by a freezing mixture, such as salt and ice, the water will freeze, or if the dish is brought into a warm room, the ice will melt. In either case, however, the *temperature* of the mixture will remain steady at 0° until either all the ice is melted or all the water is frozen.

It seems evident, then, that when ice melts, heat energy, called **latent heat**, is absorbed, which does not show itself in a rise of temperature.

196. How much heat to melt 1 gram of ice? In solving this problem we may apply the method of mixtures which was used in determining the specific heat of a metal.

If we put 200 grams of ice at 0° C into 300 grams of water at 70° C and stir them thoroughly, the temperature of the water, after the ice is all melted, will be 10° C.

Let $x =$ no. of calories required to melt 1 g. of ice.

Then $200\,x =$ no. of calories required to melt 200 g. of ice.

Also $200 \times 10 =$ no. of calories required to raise melted ice from 0° to 10°,

and $300\,(70 - 10) =$ no. of calories given out by the water in cooling.

Then $$200\,x + 200 \times 10 = 300\,(70 - 10),$$

whence $$x = 80 \text{ calories}.$$

The best experiments that have been made show that the latent heat of melting ice is just about 80 calories, which means that 80 calories are absorbed in changing 1 gram of ice at 0° C into water at 0° C.

197. Heat given out when water freezes. We have just seen that heat energy is required to pull apart the molecules of the solid ice and change it into the liquid state, where we believe the molecules are held together less intimately. Now we want to show that in the reverse process, that is, in freezing, this energy appears again as heat. We may show that freezing is a heat-evolving process in the following experiment.

FIG. 170. — Freezing water evolves heat.

If we repeat the experiment described in section 192, except that we keep the water, thermometer, and test tube (Fig. 170) very quiet, we shall be surprised to find that the water will cool several degrees below 0° C before the freezing begins. When once started by stirring or dropping in a crystal of ice, the crystals of ice form rapidly, but the temperature jumps to 0° C and remains stationary until all the water is frozen, even though the freezing mixture in the jar outside the test tube is as cool as −10° C.

People sometimes make use of the heat given out by water when it freezes, by putting pails or tubs of water in a greenhouse or a cellar to prevent the freezing of the plants or vegetables. As the water begins to freeze first, the heat evolved in the process prevents the temperature from falling much below 0° C. When a large lake freezes, the heat evolved helps to keep the temperature in its vicinity from falling as low as it does farther away.

PROBLEMS

1. How many calories of heat are required to melt 20 grams of ice at 0° C?

2. How much heat is evolved in cooling and freezing 12 grams of water originally at 10° C?

3. How many B. t. u. are required to melt one pound of ice at 0° C?

4. How much water at 100° C will be needed to melt 300 grams of snow at 0° C, and raise its temperature to 20° C?

5. If a 500-gram iron weight is heated to 250° C and placed on a block of ice, how many grams of the ice will be melted?

198. Process of boiling water. Let us fill a round-bottomed flask (Fig. 171) half full of water and put through the stopper a thermometer, an open manometer, and an outlet tube for the steam. At first, as the water is heated, the air, which is dissolved in the water, rises to the surface in little bubbles. Then bubbles of steam form at the bottom, but these collapse when they strike the upper, cooler layers of water, and disappear, causing the rattling noise known as "singing" or "simmering." When the bubbles of steam begin to reach the surface, the water is said to "boil." It will be noticed that the steam in the flask is as clear as air, but as it leaves the outlet tube it condenses and forms a white cloud or mist.

FIG. 171. — Boiling water.

As soon as boiling begins, the thermometer, which has been rising rapidly, reaches 100° C and remains stationary.

If we partly close the outlet valve, the manometer will show an increase of pressure, while the thermometer will show a rise in the temperature of the boiling water.

Finally if we remove the burner, and let the water cool a bit, we may connect the outlet tube with an aspirator, which will reduce the pressure and make the water boil again.

The process of boiling consists in the formation in a liquid of bubbles of vapor, which rise to the surface and escape. The temperature at which this takes place is the *boiling point* of the liquid.

There is a second and more exact definition of the boiling point. It is evident that a bubble of water vapor can exist within the liquid only when the pressure exerted outward by the vapor within the bubble is at least equal to the atmospheric pressure pushing down on the surface of the liquid.

For if the pressure within the bubble were less than the outside pressure, the bubble would immediately collapse. Now the pressure that would exist inside a bubble, if it could form at all, would be different at different temperatures. It is called the **vapor pressure**, or **vapor tension**, of the liquid, and we shall soon see how to determine its values at different temperatures. *The* **boiling point** *of a liquid may therefore be defined as the temperature at which its vapor pressure is one atmosphere.*

199. Effect of changing pressure. We have just seen in the experiment about boiling that if the pressure on the surface of the liquid is increased, the temperature has to be raised before the liquid will boil. If the pressure is decreased, the liquid will boil at a lower temperature. We can understand this if we recall that ordinarily the atmosphere is exerting a pressure of about 15 pounds per square inch on the surface of the liquid. If we reduce this pressure, it is easier for the bubbles of vapor to form; if the pressure is increased, it is more difficult for the bubbles to form. In any case, they will form only when the temperature is high enough so that, when they have formed, the pressure in them is equal to the pressure on the surface of the liquid. So by observing the temperatures at which a liquid boils under different pressures, we can determine how the vapor pressure of the liquid changes with temperature. Experiments have shown that, near 100° C, the vapor pressure of water increases by about 27 millimeters of mercury for each Centigrade degree rise of temperature.

Benjamin Franklin devised the following interesting experiment to show water boiling under reduced pressure.

Let a flask half full of water, which is boiling vigorously, be removed from the flame and instantly corked air-tight with a rubber stopper. We may then invert the flask, as shown in figure 172, and cool the top by pouring on cold water. The water in the flask immediately begins to boil again. This is because the steam in the top of the flask is condensed and so the pressure on the surface of the liquid is much reduced.

Sometimes it is very desirable to boil liquids at as low a temperature as possible. For example, the water is boiled away from sirup and from milk in what are called **vacuum pans**, which are merely closed kettles with part of the air pumped out. The water boils away at about 70° C and leaves the granulated sugar or milk condensed, but not cooked.

FIG. 172. — Boiling water under reduced pressure.

On the tops of high mountains the temperature of boiling water is so low that eggs cannot be cooked. In Cripple Creek, Col., about 10,000 feet above sea level, it takes about twice as long to cook potatoes as in Boston. In some high altitudes closed vessels provided with safety valves, called "digesters" or "**pressure cookers**" (Fig. 173), have to be used in cooking. Digesters are also used for extracting gelatine from bones. The effect of the increased pressure in a digester or pressure cooker is the same as in a boiler. The water in a boiler whose gauge reads 100 pounds is boiling, not at 100° C, but at 170° C or 338° F.

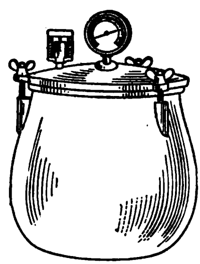


FIG. 173. — Pressure cooker.

Since we have defined the 100° point on the Centigrade scale as the temperature of boiling water, and since the temperature at which water boils is so much affected by changes in pressure, it is necessary to fix on some **standard pressure** at which thermometers are to be "calibrated" or marked. By common agreement, this standard pressure is the pressure exerted by a column of mercury 760 millimeters high, the temperature of the mercury being 0° C. The temperature at which water boils under this pressure is, by definition, 100° C.

200. Summary. What has been said about the process of boiling can be summarized as follows: —

(1) *A liquid will boil only when its temperature is such that its vapor pressure is equal to the pressure on its surface.*

(2) *What is called "the boiling point" of a liquid is the temperature at which it will boil under atmospheric pressure; that is, the temperature at which its vapor pressure is one atmosphere, or 760 millimeters of mercury.*

(3) *Every liquid has its own boiling point. The boiling point of water is by definition 100° C.*

(4) *The rule about boiling under other pressures than one atmosphere is, the higher the pressure, the higher the temperature required to make the liquid boil.*

TABLE OF BOILING POINTS

(At a pressure of 760 millimeters)

Zinc	918° C	Alcohol	78° C
Sulphur	445° C	Ether	35° C
Mercury	357° C	Ammonia	− 34° C
Saturated salt solution	108° C	Oxygen	− 183° C
Water	100° C	Hydrogen	− 253° C

201. Distillation. In many localities the only way to be sure of getting pure water is by what is called distillation.

Let us set up a boiler B and a condenser C as shown in figure 174, and color the water in the boiler with blue vitriol (copper sulphate). When the solution is boiled, the vapor or steam given off is condensed, by the continual circulation of cold water through the jacket, as a colorless, tasteless liquid, pure or *distilled water.* The non-volatile impurities, including the vitriol, are left behind in the boiler.

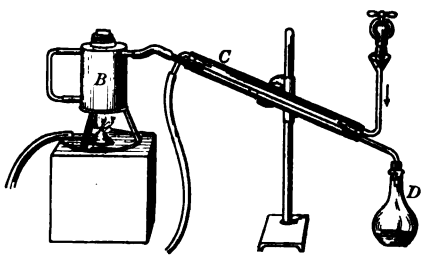


FIG. 174.—Purification of water by distillation.

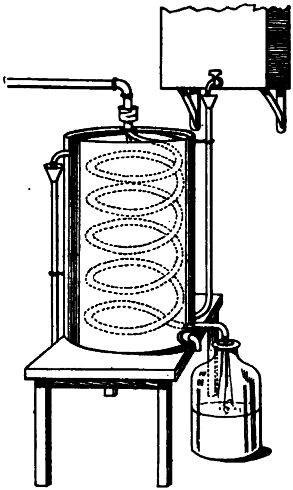


FIG. 175.—Worm condenser.

The process of **distillation** *consists of boiling a liquid and condensing its vapor.* In commercial work this is usually done in a "worm condenser." This consists of a pipe coiled into a spiral and surrounded by circulating cold water (Fig 175). In this way a large condensing surface is obtained in a small space.

When a mixture of two liquids is distilled, the liquid with the lower boiling point vaporizes and is condensed first. It can thus be separated from the one with the higher boiling point. Thus alcohol is distilled from fermented liquors by this process of **fractional distillation.** It is in this way that gasolene and kerosene are got from crude petroleum.

PROBLEMS AND QUESTIONS

1. How is the temperature of boiling water affected by taking the water to the bottom of a deep mine?
2. If water boils at 99° C, what is the atmospheric pressure?
3. If water boils at 208° F, what does the barometer read?
4. An elevation of 900 feet makes a difference of about 1 inch in the barometer. At what temperature would water boil 1500 feet above the sea?
5. What effect does salt or sugar have on the boiling point of water? Try it.
6. In distilling a mixture of alcohol and water, which liquid begins to distill over first?
7. How could you obtain fresh water from sea water?
8. Mark Twain in his "Tramp Abroad" tells of stopping on his way up a mountain to "boil his thermometer." What did he do, and why?

202. Latent heat: water to steam. When a kettle of water is put on a stove, it gets hotter and hotter until it boils. Then no matter how much heat we apply to the kettle, if there is a free outlet for the steam to escape, the temperature remains constant at 100° C or 212° F. The heat energy which seems to disappear in boiling the water is called the **latent heat of steam** or the latent heat of vaporization. When steam flows from a steam pipe into a radiator in a room, some of it condenses and gives back the heat which apparently disappeared when the water changed into steam. This latent (or hidden) heat is now understood to represent the energy needed to pull the molecules of water away from each other and set them free as steam.

203. How much heat is needed to make a gram of steam? When we want to determine the amount of heat needed to change a gram of water at 100° C into steam at 100° C, we usually apply the method of mixtures. In practice we generally try to determine the heat evolved in condensing a gram of steam by running dry steam into a given quantity of water at a known temperature for some time. We measure the rise in temperature and the increase in weight, which is the weight of the condensed steam. Then we make an equation in which the number of calories received by the water in being warmed is put equal to the calories given out by the steam in condensing to water at 100° C and by this hot water in cooling from 100° C to the temperature of the mixture.

Suppose we take 400 grams of water at 5° C and run in 20 grams of steam at 100° C, which raises the temperature of the water to 35° C. What is the number of calories of heat given out by 1 gram of steam in condensing to water at 100° C?

Let x = latent heat of steam.
Since $400(35-5)$ = heat absorbed by cold water,
and $20\,x$ = heat given out by condensing of steam,
and $20(100-35)$ = heat given out by water in cooling from 100° to 35° C,
then $400(35-5) = 20\,x + 20(100-35)$,
and $x = 535$ calories.

Recent experiments have shown that the latent heat of steam is about **540 calories**. In other words, it takes more than five times as much heat to change any quantity of water into steam as to raise the same quantity of water from the freezing to the boiling point. In English units it requires 540×1.8 or 972 B. t. u. to change a pound of water at 212° F into steam at 212° F.

PROBLEMS

1. Find the number of calories required to change 15 grams of water at 100° C into steam.

2. Compute the heat evolved by condensing 10 grams of steam at 100° C and cooling it down to 50° C.

3. How much heat will be required to convert 1 kilogram of ice at 0° C into steam at 100° C?

4. How much steam at 100° C must be run into 500 grams of water at 10° to raise it to 40°?

5. In the illustrative example in section 203, the latent heat came out 535, which is a little too low. This shows that the temperature of the mixture (35° C) was not acccurately observed. What should it have been?

6. How many pounds of coal will be needed in a boiler whose efficiency is 65%, to convert 100 pounds of water at 50° F into steam at 212° F? Assume that the heat value of the coal is 14,500 B. t. u. per pound.

204. Evaporation. Everybody is familiar with the fact that water left in an open dish gradually disappears or evaporates. **Evaporation** is different from **boiling**, in that evaporation takes place at any temperature but only at the surface of a liquid, while boiling goes on inside the liquid but only at a fixed or definite temperature. Evaporation goes on more rapidly the warmer and drier the surrounding air is. For example, wet clothes dry more quickly on a hot day than on a cold, foggy day.

205. Cooling by evaporation. If one pours a few drops of alcohol or ether on his hand, the liquid quickly evaporates, causing a marked sensation of cold. Whenever a liquid evaporates, it must get heat from somewhere, and so the

temperature of the liquid itself and of anything near it drops That is to say, heat is absorbed in the process of evaporation. It is always more comfortable on a hot day to ride in a car than to sit still, because the rapid circulation of the air makes the moisture of the skin evaporate more rapidly. This is why one can tell the direction of the wind by lifting a moistened finger; the wind blows from the side which feels cool.

206. Moisture in the air. In the summer time a pitcher of ice water is usually covered with little drops of water or "sweat." It might at first be thought that these were due to the water oozing through the pores in the side of the pitcher; but the microscope does not show any pores in glazed porcelain or glass, so we must conclude that the drops come from the surrounding air. The air is cooled by coming in contact with the cold pitcher and deposits some of its moisture. If we put a little water in a bottle and cork it tightly, the water does not evaporate because the air above the water quickly becomes "saturated" with moisture. Thus we see that air can take up only a definite quantity of moisture, depending on the temperature. This can be better understood from the following experiment.

Let us place a little water in a thin-walled flask and cork it. If we place the cask in a warm place until it becomes warm, and then cool it, its walls become dim, due to the drops of water. The warm saturated air becomes "supersaturated" on cooling.

Careful experiments show that a cubic meter of **saturated air** contains at different temperatures the following amounts of water vapor: —

2 grams at	−10° C.
5 grams at	0° C.
9 grams at	10° C.
17 grams at	20° C.
30 grams at	30° C.
597 grams at	100° C.

From this table it will be seen that air, which is saturated at one temperature, can, at a higher temperature, take up still more water vapor before becoming saturated; but if cooled, it must deposit some of the water vapor which it already has.

207. Relative humidity. Usually the air does not contain all the moisture which it can hold; that is, it is not saturated. If, however, the temperature suddenly drops, the same actual amount of moisture will saturate the air.

For example, if the water in a polished nickel-plated cup is cooled with ice below the temperature of the room, a mist will appear on the outside of the beaker. The temperature of the water when this occurs is the "dew point."

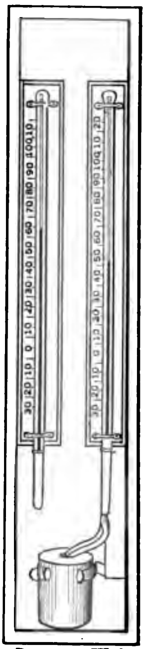


FIG. 176.—Wet and dry bulb thermometers.

The **dew point** is the temperature at which the water vapor in the air begins to condense. If the air is cooled below the dew point, some of its vapor condenses, and dew collects on objects. Thus we see that the words "dry" or "moist," as applied to the atmosphere, have a purely relative significance. They involve a comparison between the amount of water vapor actually present, and that which the air could hold if saturated at the same temperature. The ratio of these two quantities is called the **relative humidity.** For example, we may read in the newspaper that the relative humidity is 75 %. This means that the amount of water vapor actually present in the air is 75 % of what the air might have contained at the given temperature if it had been saturated.

208. Wet and dry bulb thermometers. Let two thermometers be arranged as shown in figure 176. The bulb of the thermometer at the left is dry, while the other thermometer has its bulb wrapped with cotton cloth which is kept moist by a cup of water. If we keep the air around the thermometers circulating by an electric fan, after a

little while the wet-bulb thermometer will indicate a lower temperature than the dry-bulb thermometer. This is because of the cooling caused by the evaporation from the cotton cloth. The drier the surrounding air, the more rapid will be the evaporation, and so the greater will be the difference between the wet and dry bulb thermometers. With the aid of tables furnished by the Weather Bureau, we may determine from these thermometer readings the so-called "relative humidity" of the air.

209. Practical importance of determining humidity. It is well known that a hot day in Boston is much more uncomfortable than an equally hot day in Denver. This is because a city near the ocean, like Boston, has a higher relative humidity than a city which is inland and a mile above sea level, like Denver. When the relative humidity is high, we feel "sticky" because the perspiration of the skin does not evaporate readily. On the other hand, too little humidity is injurious. Special precautions are taken to keep the air in schools, hospitals, and private houses from getting too dry in winter, and the air in greenhouses must be kept quite damp for the growth of plants. In cotton mills it has been found that the air must be rather moist to make the spinning of yarn successful.

Since the occurrence of frost in the late spring or early fall is injurious to many crops, it is often highly important that farmers should know in the afternoon whether freezing weather during the night is to be expected. The temperature of the dew point gives a ready means of predicting how low the temperature at night will drop; for when the dew point is reached, further cooling is retarded. So if the dew point is above 40° F, it is seldom that the temperature will fall to freezing in the night.

210. Dew, fog, rain, and snow. On clear, still nights the ground radiates the heat that it has received during the daytime. The grass and leaves, which can radiate heat freely, cool rapidly and soon bring the air near them below its dew point. Then moisture condenses as dew or at lower tem-

peratures as frozen dew or **frost**. This phenomenon is exactly like the formation of drops of water on a pitcher of ice water, or on one's spectacles when he comes from the cold outdoors into a warm room. Clouds covering the sky hinder the formation of dew because they restrict radiation. If the condensation of the moisture of the air is not brought about by contact with cold solid objects at the surface of the earth, but by great masses of cold air high above the earth, **clouds** are formed and **rain** may result. **Fog** is merely clouds very near the earth.

Clouds at very high altitudes may be composed of bits of ice, but, in general, clouds are made up of minute drops of water. Like particles of fine dust, very small drops of water tend to fall, but can do so only very slowly. Sometimes they fall into warm and not yet saturated layers of air, and then they change back again into vapor. Sometimes they are carried up by ascending currents of air faster than they can fall through them, and so seem to float. For example, the cloud of steam above a locomotive stack is composed of minute drops of water and yet rises with the warm air. Clouds are not durable. They simply mark the place in the atmosphere where the process of condensation of water vapor is going on. In **rain** clouds the little particles of water come together and form drops which easily over-

FIG. 177.—Snow crystals.

come the resistance of the air and fall to the ground. If the temperature of the cloud is below 32° F, the particles of water unite to form little delicately fashioned hexagonal **snow crystals** (Fig. 177).

Snow and rain together make what the "weather man" calls "precipitation." Thus in New York there are about 150 days of rain or snow each year, and the total precipitation in a year, if it did not dry up, would cover the earth to a depth of about 3 feet.

Questions and Problems

1. A room is 3 meters high, 10 meters long, and 6 meters wide. How many grams of water will be required to saturate the air at 20° C?
2. An experiment showed that on a certain day, when the temperature was 30° C, the air contained 12 grams of water per cubic meter. What was the relative humidity?
3. How do undue dryness and undue dampness affect wooden furniture?
4. What change in the thermometer usually goes with a rising barometer?
5. What happens when a moist wind from the ocean strikes a mountain range?
6. In some hot countries the people cool their drinking water by setting it in jars of porous earthenware, in a shady place, where there is a current of air. Explain.
7. Milk used to be set away in shallow pans for the cream to rise. Now they use cylindrical tanks of small area and quite deep. Which is the better, and why?
8. Why do clothes dry best on a windy day?
9. Why does sprinkling the street on a hot day cool the air?

211. Freezing by boiling. The fact that a large quantity of heat is needed to vaporize a substance is often made use of in getting low temperatures.

If a cylinder of liquefied carbon dioxide is tilted, as shown in figure 178, and the valve is opened, the liquid released from pressure vaporizes so rapidly as to cool everything, including the rest of the liquid, and so

some of it is frozen. After the valve has been open a short time, the bag is filled with a white solid, frozen carbon dioxide. This solid evaporates very readily, and gives a temperature as low as $-80°$ C. If the solid is put in a beaker and mixed with ether, the mixture will freeze a test tube of mercury. The ether serves to carry the heat quickly from the test tube to the solid.

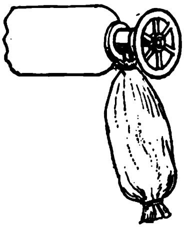


Fig. 178. — Liquid carbon dioxide being frozen.

212. Artificial ice. In the manufacture of artificial ice and in refrigerating plants (Fig. 179), gaseous ammonia is compressed by a pump and then cooled until it liquefies. During this process of compression and of condensation, heat is evolved, which is removed by passing the ammonia through a pipe covered with running water. The liquefied ammonia is then piped to the ice tank or cold-storage room, and allowed to expand through a valve with a small opening. This checks the flow, and so enables the pump to maintain enough pressure to keep the ammonia in liquid form on its way to the valve; while beyond the valve the pressure is very small, so that the ammonia expands and evaporates rapidly. While doing so, it absorbs heat from the refrigerating room. It is then ready to be compressed again.

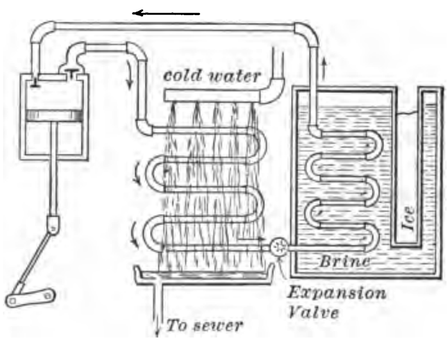


Fig. 179. — Diagram of cold-storage plant.

In the manufacture of ice, the expansion pipes pass through a brine tank in which are smaller tanks of pure water. When the water in these tanks is frozen, the tanks are pulled up and the ice removed and stored. The ammonia is used

over and over again, but power must be constantly supplied to keep the compressor working.

SUMMARY OF PRINCIPLES IN CHAPTER XI

Heat units: —

1 B. t. u. = heat to raise 1 lb. of water 1° F.

1 calorie = heat to raise 1 gram of water 1° C.

Specific heat = calories to raise 1 gram of substance 1° C.

Specific heat of water = 1.

Method of mixtures: —

Heat given up by hot bodies = heat absorbed by cold bodies.

Pressure: —

***Lowers* freezing point of water 0.0072° C per atmosphere.**

***Raises* boiling point of water 0.037° C per millimeter of mercury.**

Latent heat of melting = heat absorbed during melting,
= heat yielded during freezing.

Value for water, 80 calories.

Latent heat of vaporization = heat absorbed during evaporation,
= heat yielded during condensation.

Value for water, 540 calories.

Relative humidity

$$= \frac{\textbf{actual moisture in air}}{\textbf{moisture sufficient to saturate air at same temp.}}$$

QUESTIONS

1. If you know the dew point to be 10° C, how could you find the relative humidity at 20° C?

2. Human hair when treated with ether is very sensitive to moisture. When it is moist it contracts, and when it dries it elongates. Explain how a moisture gauge or "hygrometer" could be made with a hair.

3. Why do they not cast gold money instead of stamping it with a die?

4. Why is a burn from live steam so severe?

5. Why does one sometimes "catch cold" by sitting in a draft of cool air after taking violent exercise?

6. How low may the temperature fall during a rain?

7. Why can mercury mixed with zinc and tin be purified by distillation?

8. Why is it difficult to make snowballs out of dry snow?

CHAPTER XII

HEAT ENGINES

The invention of the steam engine — boilers — slide-valve and Corliss engines — expansion — compounding — condensers — efficiency — steam turbines — 2-cycle and 4-cycle gas engines — balance sheets of engines — mechanical equivalent of heat.

213. The invention of the steam engine. In our age no other machine is of such importance as the **steam engine.** It furnishes the driving power for running a countless number of machines in our shops and factories, as well as for transportation on land and sea.

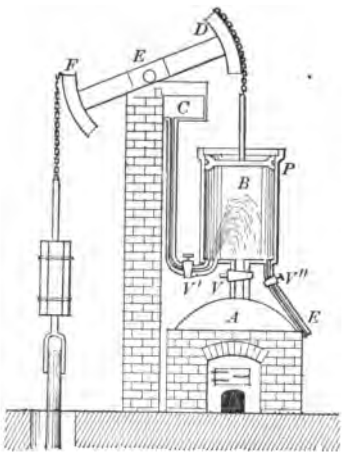


Fig. 180. — Newcomen's steam engine.

Up to about two hundred years ago steam had been used only in various devices, called steam fountains, for raising water. In 1705 the first successful attempt to combine the ideas of these devices into an economical and convenient machine was made by **Thomas Newcomen** (1663–1729), a blacksmith of Dartmouth, England. This machine was called an "atmospheric steam engine" (Fig. 180). It consisted of a boiler *A*, in which the steam was generated, and a cylinder *B*, in which a piston moved. When the valve *V* was opened, the steam pushed up the piston *P*. At the top of the stroke, the valve

V was closed, the valve V' was opened, and a jet of cold water from the tank C was injected into the cylinder, thus condensing the steam and reducing the pressure under the piston. The atmospheric pressure above then pushed the piston down again.

This machine was used to pump water from mines. It consumed a great deal of fuel, because the cold water cooled the cylinder walls so much that when the steam was turned in, much steam condensed before the piston was raised.

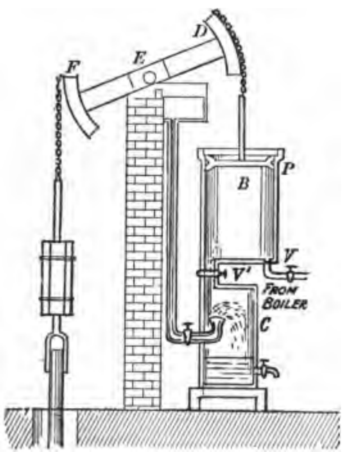


FIG. 181.—Watt added a separate condenser.

The next great step in the development of the steam engine came through a Scotch instrument maker, **James Watt** (1736–1819). He arranged a separate vessel for condensing the steam, as shown in figure 181. This condenser, C, was connected with the cylinder through a valve V'. When the piston had reached the top of the cylinder, the valve V was closed and V' was opened. Then the steam rushed from the cylinder into the condenser, which was kept cold and under less than atmospheric pressure. At first these valves V and V' had to be operated by hand, but later, it is said, a boy named Potter, whose job it was to turn these valves, connected the valve handles by cords to the beam ED in such a way that the machine became automatic.

In all these crude machines the steam simply furnished the vacuum, and atmospheric pressure did the work. Later, Watt made a machine with a closed cylinder and a piston that was pushed down as well as up by steam. By the use

of a connecting rod and crank shaft, he contrived to change the back-and-forth motion of the piston to a rotary motion, and so made the steam engine available for many new uses. Within a few years the development of the steam engine revolutionized most lines of industry.

214. A modern steam plant. In a modern steam plant the steam is made in a boiler, is used in a steam engine, and is got rid of in an exhaust or condenser. We will discuss these in turn.

215. Steam boilers. A **fire-tube boiler** consists of a steel cylinder which sometimes stands on end, as in the small "donkey engines" used with derricks, but generally is set on its side,

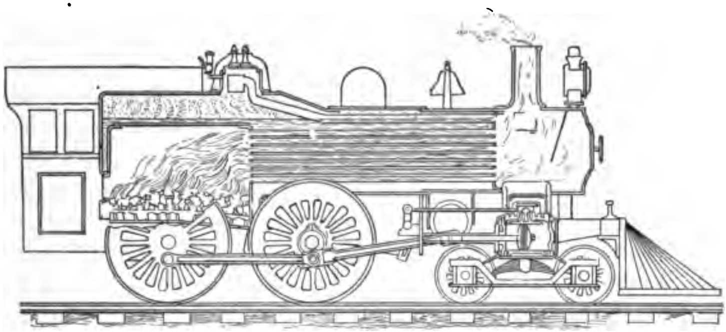


FIG. 182. — Section of a locomotive boiler.

as in locomotives (Fig. 182). Running through this cylinder are tubes, three or four inches in diameter, through which the fire and smoke pass. The water and steam fill the rest of the cylinder outside the tubes. These tubes give the boiler a much greater heating surface, so that it makes more steam per hour. Such boilers are called **fire-tube** boilers. In another type, called a **water-tube** boiler (Fig. 183), the water is inside the tubes and the fire is outside. Such a boiler consists of a large number of tubes, inclined at an angle and fastened at each end into vertical "headers"; these headers communi-

cate with a drum above, which is half full of water, the remainder of the drum forming a space for steam. The water descends by the back headers, rises through the inclined tubes, and passes up the front headers, thus maintaining a very good circulation. The fire is placed under the front end of the tubes; the gases are deflected by brick walls, so that they pass completely over and under the whole length of the tubes. In some water-tube boilers the fire grate is sloping and arranged like a flight of steps. The coal is automatically fed through a chute from the coal loft above to the grate. The principal advantages of this type of boiler are its great freedom from risk of explosion and its ability to make steam quickly. A modified form of this boiler is generally used in marine work.

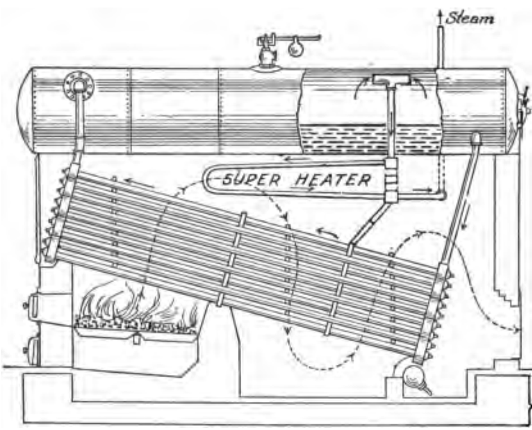


FIG. 183. — Section of water-tube boiler.

Not only is it desirable to get the greatest quantity of steam with the least expenditure of fuel, but it is also essential to keep the steam pressure constant and to prevent an explosion which may have frightful consequences. Therefore every boiler is equipped with a **steam gauge**, which is merely a Bourdon pressure gauge (section 75), and a **water gauge** (section 62), which enable the engineer in charge to watch the pressure and water level in the boiler. If the water level is too low, there is danger of burning the tubes

and plates and perhaps of wrecking the boiler; if it is too high, water is liable to be carried along with the steam and so damage the engine. Besides these devices, every boiler must have a **safety valve**, which automatically lets the steam blow off when the pressure exceeds a certain limit. A simple form of safety valve is shown in figure 184. In some forms a spring is set so as to release the steam if the steam pressure becomes too great inside the boiler.

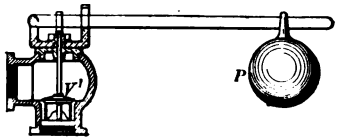


FIG. 184. — Safety valve.

In order to make steam rapidly, the fire must burn fiercely, which requires a good draft. To get this, tall chimneys are sometimes used, and at other times a forced draft is made by a big fan. On battleships a forced draft is often obtained by making the whole fireroom, within which the stokers work, air-tight, and keeping it full of air under pressure, supplied by blowers or pumps as fast as it can escape through the fires.

One pound of coal, whose heat value is 14,000 B. t. u., could change 14.4 pounds of water at 212° F into steam at 212° F if no heat were wasted. In actual practice, one pound of coal evaporates between 8 and 10 pounds of water "from and at 212° F," which means an efficiency of from 55 to 70 %. One great source of loss of heat is the flue gases. Smoke pouring from the chimney means that just so much unconsumed fuel is going to waste, and, what is worse, is adding to the dirty atmosphere of the neighborhood. To-day steam engineers are able to design boilers which, when properly stoked, produce no smoke.

216. Steam engine. The type of engine most commonly used for small plants and for locomotives is the **slide-valve engine** (Fig. 185). Steam comes from the boiler into a box or steam chest, and then into the working end of the cylinder

through a passage shown by the arrows at the right of the picture. At the same time the spent steam in the other end of the cylinder is escaping through the hollow interior of the valve, to the exhaust passage. It then escapes to the air, or to the condenser, through a pipe at the back, which does not show in the figure. At the end of the stroke the valve is pulled far enough to the right to admit live steam to the left-hand end of the cylinder, while the spent steam in the right-hand end escapes into the exhaust.

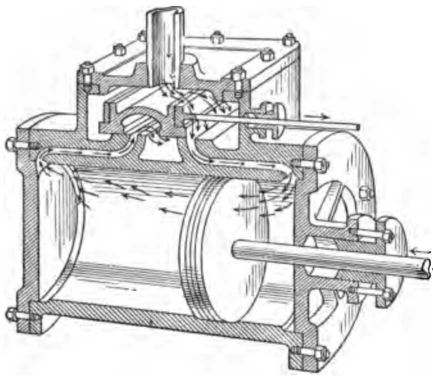


Fig. 185. — Slide-valve steam engine.

In large steam engines **Corliss valves** are more often used. A Corliss valve (Fig. 186) opens and closes by *turning* a little in its seat. In a Corliss engine there are four such valves — two at each end of the cylinder. Two of them, *A* and *B*, are for admitting the steam, and two, *C* and *D*, for letting the steam out. When valve *B* is open to admit steam, valve *D* is also open to let steam out of the other end of the cylinder, while *A* and *C* are closed; on the reverse stroke, *A* and *C* are open, while *B* and *D* are closed. These valves are automatically opened and closed at the proper time by the engine itself. The fact

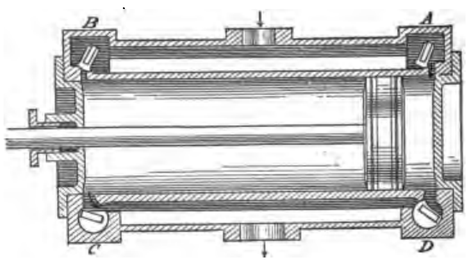


Fig. 186. — Corliss steam engine.

that the time at which each valve opens can be accurately adjusted independently of the other valves makes Corliss engines more efficient than slide-valve engines, and has led to their extensive use in large installations.

217. Expanding steam. If live steam from the boiler is allowed to push the piston through its entire stroke, and is then thrown away, that is, allowed to pass into the atmosphere or into a condenser, it is evident that much energy is wasted. To get more work out of the steam, the valve is closed after the piston has made about $\frac{1}{4}$ or $\frac{1}{8}$ of its stroke, and the steam is allowed to **expand** through the rest of the stroke. The pressure continues to drop after the "cut-off," as shown in figure 187, where the pressure P is represented vertically and the stroke horizontally. Such pressure diagrams can be made automatically by the engine itself while in actual operation, and enable those in charge to adjust the valves properly.

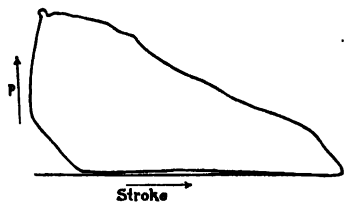


FIG. 187. — Pressure in cylinder of engine.

218. Compound engine. Another device for getting more work out of the steam is to use the steam at high pressure in one cylinder, then allow it to pass into a second, larger cylinder, where it expands some more, and sometimes into a third and a fourth cylinder. These are called **compound, triple** and **quadruple expansion** engines. When the expansion and consequent cooling of the steam take place in steps, there is no large drop in temperature in any one cylinder. So the walls of a cylinder never get much cooler than the incoming steam, and there is little condensation in the cylinders. In a simple engine, the initial steam pressure varies from 80 to 100 pounds, while in compound engines the initial pressure is usually higher, from 100 to 175 pounds. A simple engine requires from 17 to 35 pounds of steam an hour for each

horse power developed, while a compound engine may need as little as 11.2 pounds of steam per horse-power hour. Triple-expansion engines are usually used in marine work.

219. Condenser. When its exhaust pipe opens directly into the atmosphere, an engine is called a **non-condensing** engine. The power depends on the excess of the steam pressure in the boiler above that of the atmosphere outside. Ordinary locomotives and most small engines are of this type. In fact the locomotive depends on the escaping steam to furnish a draft for the boiler.

Greater economy is obtained by sending the exhaust steam to a vacuum chamber, or **condenser.** In one type the steam coming from the engine is condensed by a jet of cold water, and in another type it is condensed in tubes surrounded by cold water. A small pump is used to pump out the condensed steam as well as any air which may have leaked in. Such engines are known as **condensing** engines. Marine engines are always condensing engines.

220. Efficiency of a steam plant. We have already seen that the modern steam boiler has an efficiency of about 70%, but there are still larger losses in the engine itself. The escaping steam from an engine always carries away a large amount of unutilized heat energy. It can indeed be proved that the greatest efficiency possible for a steam engine is represented by the fraction

$$\frac{T_1 - T_2}{T_1},$$

where T_1 is temperature (Absolute) of the steam supplied and T_2 is temperature (Absolute) of the steam rejected.

For example, an engine running at 163 pounds boiler pressure takes in steam at about 185° C, and T_1 is 458°. If the temperature of the exhaust steam is 100° C, T_2 is 373°. Such an engine cannot possibly have an efficiency greater than

$$\frac{458 - 373}{458} = 18.5\%.$$

For this reason steam engineers try to use high-pressure steam, because of its high temperature, so as to make $(T_1 - T_2)$ as large as possible. The temperature T_1 is sometimes still further increased by passing the steam through pipes (Fig. 183) in the furnace to "superheat" it.

It must be remembered that this 18.5 % is the efficiency of the engine alone, so that the efficiency of the engine and boiler would be 18.5 % of 70 %, or only about 13 %. This means that about 87 % of the energy of the coal would not be converted into mechanical energy. By using very high temperatures, the latest style of quadruple expansion and condensing engine has been made to utilize about 20 % of the energy originally in the coal. The ordinary locomotive, however, does not utilize more than 8 %.

PROBLEMS

1. The area of the piston of a steam engine is 120 square inches and its stroke is 2 feet. If the "mean effective pressure" of the steam is 50 pounds per square inch, what is the total force exerted on the piston?

2. In problem 1, how many foot pounds of work are done in one revolution of the shaft (two strokes)?

3. If the engine in problem 1 is making 150 revolutions per minute, what is its "indicated horse power"; that is, what is the rate in H. P. at which the steam does work on the piston?

4. A locomotive with cylinders 18 inches in diameter and a stroke of 2 feet is provided with driving wheels 6 feet in diameter. If the mean effective pressure of the steam in the cylinder is 60 pounds per square inch, and the engine is making 50 miles an hour, what is the indicated horse power?

5. How much mean effective steam pressure will be needed to get 10 horse power from a "donkey engine" running at 200 revolutions per minute? (Assume area of piston to be 50 square inches, and stroke 1 foot.)

221. Steam turbine. Thus far we have been describing **reciprocating** engines, in which the back-and-forth motion of the piston rod is turned into rotary motion by means of a crank and connecting rod. Since the piston must come to a standstill at the end of each stroke, this means in high-

speed engines very frequent starting and stopping, which causes so much shaking as to require big and expensive foundations. On steamships the continual jarring causes a disagreeable vibration. A new and distinctly different type of engine called a **steam turbine** has been developed in recent years in which there is no reciprocating motion.

222. Curtis turbine. Steam turbines can be divided into two main classes, of which the Parsons and the Curtis turbines are typical representatives. The **Curtis** turbine is, in principle, like the Pelton water wheel (sections 78 and 79). Steam is delivered to the machine through nozzles in which it expands and gains a high velocity. It then strikes against blades fastened to the edge of a revolving disk and gives up its kinetic energy to them. In some forms of turbine (Fig. 188) there is only one set of nozzles, and the steam expands in one step from the boiler pressure to the condenser vacuum. Under such conditions the speed of the steam as it strikes the blades is so great, often more than 4000 feet per second or 2700 miles per hour, that it is difficult to handle it efficiently. Curtis turbines are therefore built in from three to six sections, each section being a complete turbine with its nozzles and wheel, and the steam is run through the sections in succession, as in a compound or multiple-expansion engine.

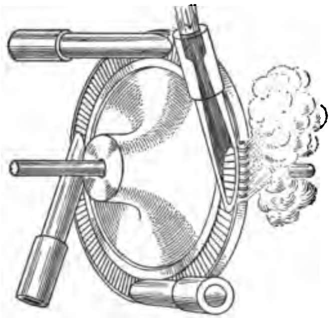


FIG. 188. — Steam turbine with one set of nozzles.

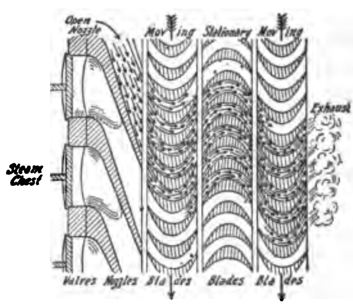
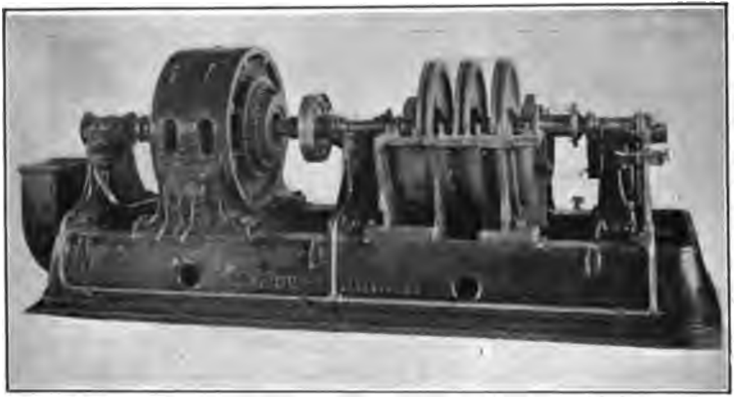
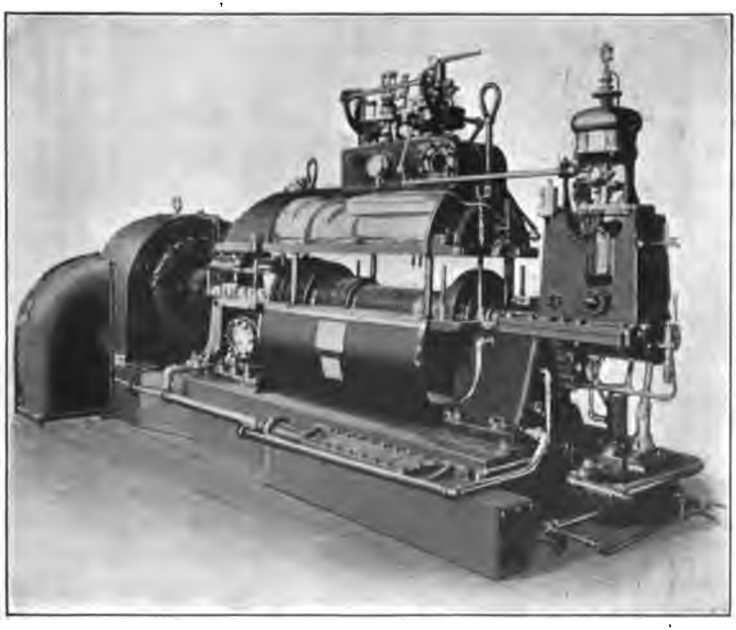


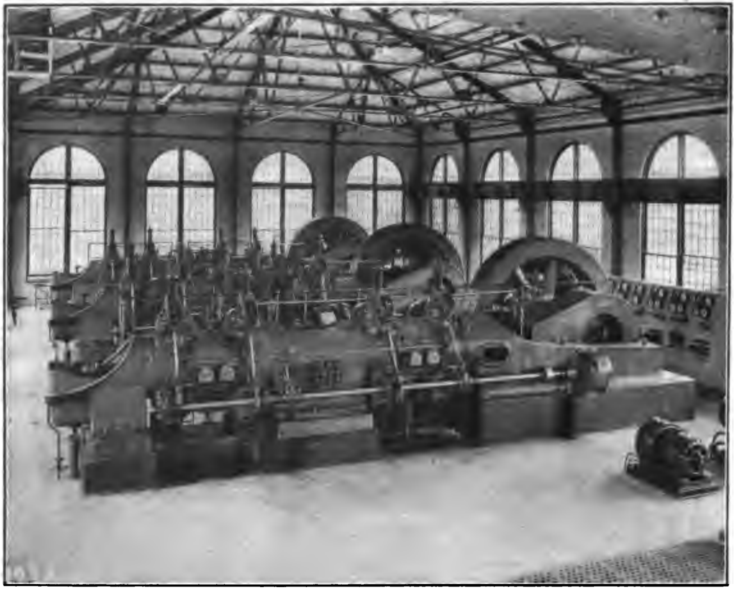
FIG. 189. — Moving and stationary blades in a Curtis turbine.



A Curtis turbine (at the right) with the upper half of its casing removed. There are three wheels and two rows of blades on each wheel. It is used to drive the generator at the left.



A Westinghouse turbine with the upper half of its casing lifted. There are a great many rows of moving blades. The balancing dummies are at the near end. The generator is at the other end, and is cooled by air drawn in through the duct.



Three large gas engines run by blast-furnace gas. Each engine has two cylinders set end to end, the two pistons being on one long piston rod running through both ends of each cylinder. The half-time shaft that works the valves is plainly visible.

FIG. 190. — Partly completed wheel, with three rows of blades, from a Curtis turbine.

In order to reduce still further the speed at which the blade wheels have to run, they are so designed that each jet has two or three chances at a given blade wheel before losing all its velocity. As it escapes at reduced speed from each set of moving blades, it is caught by guides attached to the surrounding casing and turned around so as to strike another set of moving blades on the same wheel, as shown in figures 189 and 190.

223. Parsons turbine. The **Parsons** turbine is somewhat like a succession of windmills set in line behind each other. The steam flows along the turbine from one end to the other in the annular space between the cylindrical drum or rotor and a slightly larger cylindrical casing, and acts on the windmill-like blades fastened to the drum. In passing through these rows of moving vanes, the steam would quickly get to spinning with the rotor and would then fail to act effectively on the later vanes, if it were not for the rows of stationary guide blades attached to the casing. These project between the rows of moving blades, catch the steam as it comes through, and direct it against the next row of moving blades at the proper angle. Thus the steam goes zigzagging down the annular space, striking first a row of fixed blades, then a row of moving blades, then another row of fixed blades, and so on. As the steam flows along, its pressure decreases and it expands; so the space between the rotor or drum and the outer case has to increase gradually as the low-pressure end is approached, to give the steam the extra space it requires. This is done by making the blades short at the inlet end and long at the outlet end, and by occasionally increasing the diameter of both rotor and casing, as shown in figure 191.

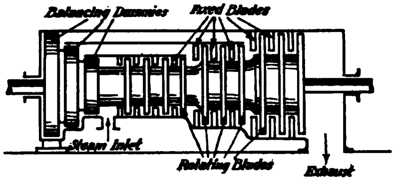


FIG. 191. — Section of a Parsons turbine.

224. Advantages of turbines. Turbine engines always run at high speed, so a large amount of power can be delivered by a small machine. This makes them especially valuable in city power stations where land is expensive. Furthermore, their lightness and steadiness make smaller and cheaper foundations sufficient. They have been installed also on large, high-speed passenger vessels and on torpedo boats and destroyers. To be operated most efficiently they should work through a wide range of temperature, corresponding to a boiler pressure of from 200 to 250 pounds, and a very perfect condenser vacuum (often better than 29 inches). A large supply of cool condensing water is therefore desirable, and turbines are especially adapted for power stations on rivers, lakes, or the ocean. Under such conditions, when working at their maximum capacity, they are slightly more efficient than even the best reciprocating engines.

225. Gas engine. The essential difference between a steam engine and a gas engine is that in the steam engine the fuel is burned under a boiler and the working substance, steam, is conducted to the engine in pipes, while in the gas engine the fuel is burned in the cylinder of the engine and the hot products of combustion are themselves the working substance. In other words, the gas engine is an **internal combustion** engine. The fuel, gasolene, is a liquid which is converted into a gas in what is called **a carbureter.** The liquid fuel is sprayed into the carbureter, vaporizes, and is mixed with the proper amount of air. This mixture of gas and air is compressed in the cylinder of the engine and then exploded by an electric spark, which causes the exceedingly rapid burning of the gas. This results in an enormous increase in pressure, which pushes out the piston. Then the exploded gases are forced out of the cylinder and a new charge of gas and air are taken in.

Inasmuch as the cylinder has to be a furnace as well as a cylinder, it would get dangerously hot if it were not cooled

from the outside. It may be **water-cooled** by surrounding it with a jacket or outer case, in which water is circulated; or it may be **air-cooled** by giving it a corrugated outer surface which radiates heat rapidly, and forcing a stream of air against this surface.

226. Two-cycle engine. In the two-cycle gas engine, we have **one explosion for every two strokes** or for each revolution of the crank shaft. A simple form, such as is used on motor boats, is shown in figure 192. The explosive mixture is taken into an air-tight crank case and slightly compressed on the outward or down stroke of the piston. As the piston nears the bottom of its stroke, it uncovers first the **outlet port**, *E*, letting part of the spent gases in the cylinder blow off, and then the **inlet port** *B*. The slightly compressed charge in the crank case then rushes into the cylinder, sweeping out the rest of the exploded gases before it. On the upstroke of the piston the ports are covered and the fresh charge is considerably compressed. As the piston passes its upper dead center (or soon afterward) the charge is exploded, and expands at a much higher average pressure than during the compression, giving back the work of compression and considerably more besides. Such an engine is called **single acting**, meaning that work is done only on one side of the piston.

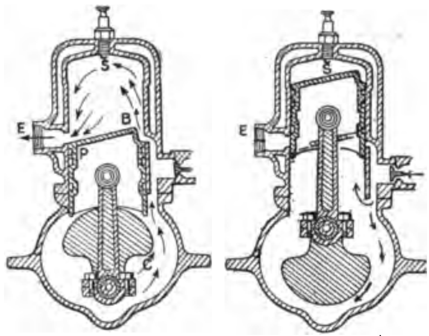


FIG. 192. — Two-cycle engine.

If the spark does not come at the upper dead center, but part way down the expansion stroke, the power yielded is much less. This is done to make a boat run slowly. Adjusting the electrical connections so as to bring the time of explosion

nearer the upper dead center is called "advancing the spark." Running on a "retarded spark" wastes gasolene, because the amount used per stroke is the same as at full power.

The only true valve in this engine is a light clap valve, where the fresh gases enter the crank case.

The disadvantage of this style of engine is that some of the fresh gas is lost with the spent gases through the exhaust, so that it uses more gasolene than some other styles. But, on the other hand, it is very simple and gives a push every revolution.

227. Four-cycle engine. In the four-cycle engine, we get a push or thrust only once in every two revolutions or every **four strokes** of the piston. Four-cycle engines, like two-cycle engines, are usually single acting.

The four-cycle type is the one most commonly used for automobiles and for stationary work. The four strokes of the piston, corresponding to two revolutions of the shaft, are shown in figure 193. It will be noticed that whereas the *two-cycle* type has *no* valves in the cylinder, the *four-cycle* has *two* valves, one for the intake of gas and air and another for the exhaust of the spent gases. These valves are operated mechanically by cams on a small **half-time shaft**, which is driven through gears at half the speed of the main shaft. In figure 193, (1), the intake valve is open, and the piston is going down, thus drawing in the explosive mixture. In (2) the return or back stroke of the piston compresses the mixture. In (3) the mixture has been ignited by an electric spark or flame, and power is obtained from the thrust of the expanding gas on the outward stroke. This is the working stroke. In (4) the exhaust valve is open and the spent gases are being pushed out of the cylinder by the returning piston. Then

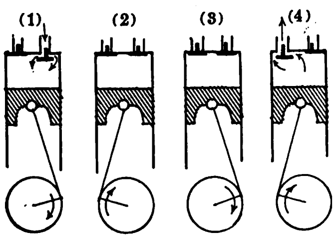


FIG. 193. — Four cycles of a gas engine.

the whole cycle is repeated again. Since power is obtained only on every alternate outward stroke, a heavy flywheel is used to keep the engine going during the other three strokes, or, as in automobiles, four such engines may act on the same shaft, so arranged that the explosions in the several cylinders take place successively, one for every half revolution of the shaft.

228. Disadvantages of the gas engine. Although the internal combustion engine has been immensely improved in the last decade, it is still a very sensitive machine. The spark must come at just the right time, and must come every time. The method for producing the spark will be described in Chapter XVII. The steam engine is more easily varied in speed than the gas engine. To be sure, it is possible to vary the speed of a gas engine somewhat by advancing or retarding the spark, and by controlling the supply of gas ; nevertheless, automobiles have to use gears to get a sufficient variety of speeds at full power. Then, too, a steam locomotive will run either way by simply shifting the slide valve, while the gasolene automobile has to use a reversing gear. Finally a gas engine is not always free from noise and smell.

229. Advantages of the gas engine. On account of lightness and compactness, and the small space occupied by the fuel, there has been a phenomenal development in the manufacture of gasolene engines for small pumping stations, shops, and factories, as well as for automobiles, launches, and aëroplanes. The gas engine does not require any stoking of a boiler or constant care to keep up the right pressure of steam. In fact, once started it requires very little attention. It can be started at a moment's notice, while if a steam engine and its boiler have been "shut down," it takes a good while to get up steam. Furthermore, no fuel is wasted when a gas engine is shut down at night or between periods of use. In efficiency the modern gas engine ranks much higher than the steam engine.

230. Balance sheet of heat engines. When an engine is tested, a heat balance sheet is usually made up. This is somewhat like a cash account, in that it accounts for all the energy delivered to the engine by the fuel. These heat balance sheets vary somewhat for different engines even of good design, but the following are fairly typical for large and efficient engines of the two types: —

Steam Engine		Gas Engine	
Useful work	15%	Useful work	25%
Friction	5%	Friction, etc.	10%
Exhaust	45%	Exhaust	30%
Up the chimney	35%	Jacket	35%
	100%		100%

231. Mechanical equivalent of heat. We have been considering the efficiency of engines without stopping to describe how it is measured. Evidently we must have some way of comparing the **output**, which would naturally be measured in foot pounds or kilogram meters, with the **input**, which would naturally be measured in B. t. u. or calories. This involves finding a definite relation between a foot pound and a B. t. u., or between a kilogram meter and a calorie. This problem was not solved until about the middle of the last century, when an Englishman, Joule (1818–1889), did his famous experiment of churning water.

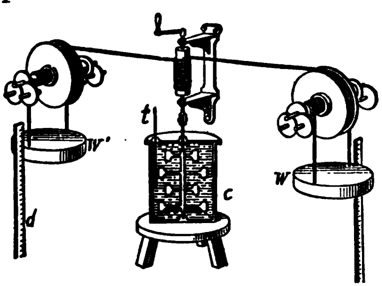


FIG. 194. — Joule's machine to find mechanical equivalent of heat.

He arranged a paddle wheel in a box of water (Fig. 194). The paddles were turned by weights which descended and thus unwound cords on the spindle of the wheel. The water was kept from following the rotat-

ing paddles by fixed paddles which projected from the sides of the box.

In this experiment the mechanical work put in could be measured by multiplying the weights by the distance through which they fell; and the heat produced could be measured by multiplying the weight of the water by the rise in temperature. Great care was taken to prevent any loss of heat. The result of this and many other experiments of a similar nature led Joule to announce this principle: *The number of units of work put in is always proportional to the number of units of heat produced.*

As a result of Joule's experiments and also of the more accurate experiments of Rowland (1848–1901) and of many others, we believe that 778 *foot pounds of work are equivalent to the heat required to raise one pound of water one degree Fahrenheit*, or that *the energy required to heat one kilogram of water one degree Centigrade is equal to the work done in raising one kilogram to a height of* 427 *meters.*

1 B. t. u. = 778 foot pounds of work.
1 kilogram calorie = 427 kilogram meters of work.

To compute the efficiency of an engine we have, therefore, to divide the work done by the heat put in, *expressing both in the same units* by means of the above relationships.

This work of Joule's was a clinching argument in favor of the principle of the conservation of energy, for it meant that heat and work are but different forms of energy.

Problems

1. If a horse power is equal to 33,000 foot pounds of work per minute, how many foot pounds are there in a horse power hour; that is, in the total amount of work produced by a 1 H. P. engine working for 1 hour?

2. A pound of average coal yields 14,500 B. t. u. when burned. To how many foot pounds is this heat equivalent?

3. From the results of problems 1 and 2, calculate the horse power hours per pound of coal.

4. A test of a certain steam engine showed that 1 pound of coal generated 1 horse power hour; from the three preceding problems compute the efficiency.

5. Calculate the efficiency of a gasolene engine from the following data: 16 cubic feet of gas were used per horse power hour; 1 cubic foot of gas yields 700 B. t. u.

SUMMARY OF PRINCIPLES IN CHAPTER XII

The mechanical equivalent of heat is the value in foot pounds of one B. t. u. or in kilogram meters of one calorie.

$$\textbf{1 B. t. u.} = \textbf{778 ft. lb.}$$
$$\textbf{1 kg. cal.} = \textbf{427 kg. meters.}$$

$$\textbf{Efficiency} = \frac{\textbf{output}}{\textbf{input}}.$$

Both must be expressed in *same unit* by means of above relations.

The conservation of energy in engines requires that all energy supplied as heat of combustion of the fuel be accounted for as useful output or specified waste ("making up the heat balance sheet of the engine").

QUESTIONS

1. Why does the steam jacket increase the efficiency of a steam engine?

2. Does the water jacket increase the efficiency of a gasolene engine?

3. What did Count Rumford learn about heat while boring cannon for the Bavarian government?

4. How are the cylinders of engines lubricated?

5. How does a ship equipped with steam turbines reverse its propellers?

6. Describe the reversing mechanism of a locomotive.

7. Is an ordinary gas engine self-starting? How are automobile engines made self-starting?

8. When you see steam coming from the exhaust pipe of a steam engine in puffs, do you know whether it is a condensing or non-condensing engine?

9. Why are condensers not used on locomotives?

10. What advantages has oil as a fuel for locomotives and steamships?

11. Why are marine engines always condensing engines?

12. How would you compute the efficiency of a gun regarded as a heat engine?

13. What makes the water circulate in a water-tube boiler?

14. Why does a high-speed turbine give more power than a low-speed reciprocating engine of about the same size?

15. What is the use of the radiator on an automobile?

CHAPTER XIII

MAGNETISM

The lodestone — magnetic poles — attraction and repulsion — the compass and magnetism of the earth — magnetic field — induced magnetism — permeability — theory of magnetism.

232. The lodestone. For many centuries it has been known that a certain kind of rock, called the lodestone, has the power of attracting iron filings and small fragments of the same rock. Its abundance near Magnesia in Asia Minor led the Greeks to call it "magnetite" or "magnetic" iron ore.

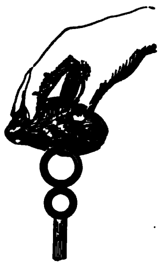


FIG. 195. — Lodestone attracts iron.

Let us take a piece of magnetite (Fe_3O_4) and show that it picks up pieces of iron (Fig. 195), but does not pick up copper or zinc. We may magnetize a knitting needle by stroking it with a piece of magnetite.

This kind of iron ore occurs in many places in this country as well as in Norway and Sweden. When a steel bar is rubbed with such a natural magnet, the steel itself becomes magnetic and is then called an artificial magnet. In a later chapter we shall learn how to make magnets by using an electric current.

233. Magnetic poles. It was a good many years before any one in Europe noticed that the magnetic property of a lodestone was concentrated more or less definitely in two or more spots, and that if a somewhat elongated lodestone with only two of these spots, and those near its ends, is hung by a thread, it will set itself with one spot toward the north and the other toward the south. We now use magnetized

needles instead of lodestones, and call such an arrangement a **compass**, and we all know how valuable it is to mariners and explorers. Probably the Chinese had compasses many years before Europeans reinvented them.

The two spots which point one to the north and one to the south are called the **poles** of the magnet; one is called the **north-seeking** pole (*N*) and the other the **south-seeking** pole (*S*).

234. Magnetic repulsion. It was many centuries after people had known that magnets would attract things before they learned that magnets sometimes repel things.

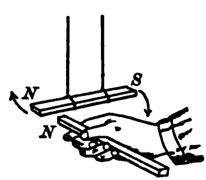


FIG. 196. — Magnetic repulsion.

If we bring the north-seeking or *N*-pole of a magnet near the *N*-pole of a suspended magnet, the poles repel each other (Fig. 196). If we bring the two *S*-poles together, they also repel each other. But if we bring an *N*-pole toward the *S*-pole of the moving magnet, or an *S*-pole to the *N*-pole, they attract each other.

That is,

Like poles repel each other,
Unlike poles attract each other.

Experiment shows that these attractive or repulsive forces vary inversely as the square of the distance between the poles.

235. Declination and dip. Soon after the compass was invented, it was noticed that it did not point true north and south. For a long time it was supposed that this deviation or **declination** was everywhere the same, until Columbus, on his way to America in 1492, discovered near the Azores a place of no declination. Evidently an exact knowledge of the declination at different places is of the greatest importance to mariners and surveyors, and so careful maps are published by the different governments giving lines of equal declination. Figure 197 shows such a map. From this map it will be observed that in the extreme eastern section of the

United States the declination is as much as 20° W. This decreases to zero at a place near Cincinnati, O., and becomes an easterly declination amounting to 20° E. in the northwest.

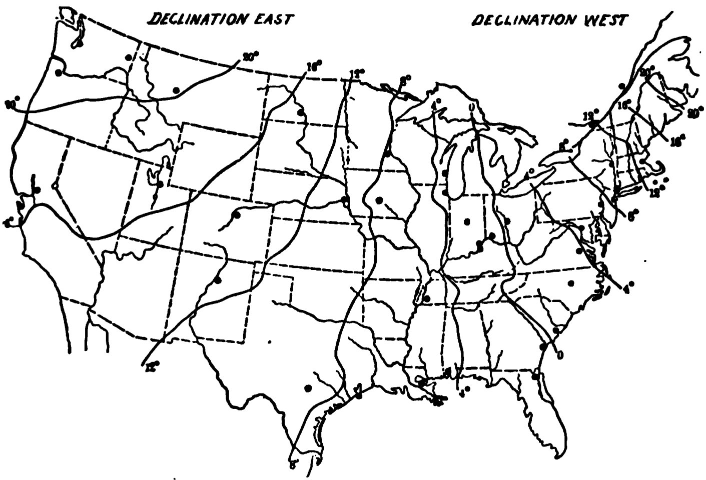


FIG. 197. — Map showing declination of the compass in the United States.

It was nearly a hundred years after Columbus' time before it was discovered that if a compass needle is perfectly balanced so that it can swing up and down as well as sidewise, its north-seeking pole will **dip** down at a considerable angle (Fig. 198). This angle increases as one goes farther north, and decreases as one goes south. Along a line near the equator there is no dip. In the southern hemisphere the north-seeking pole of a needle points up in the air, and recently Shackleton's South Polar Expedition found a point on the great Antarctic continent where a needle would hang vertically with its north-seeking pole on top.

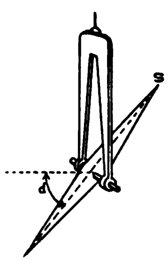


FIG. 198. — Compass needle to show magnetic dip.

236. The earth a magnet. An Englishman, Gilbert, in the sixteenth century was the first to explain these curious magnetic phenomena. He had ground a little lodestone into the shape of a globe, and noticed that when tiny compass needles were brought near it, they acted just like compasses on the surface of the earth. So he called his lodestone globe the "terrella" or "little earth" (Fig. 199), and came to believe that it gave a true representation of the earth itself.

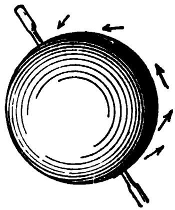


Fig. 199.—Gilbert's terrella or little earth.

The earth is, then, simply a huge magnet, much thicker in proportion to its length than the magnets with which we are familiar in laboratories, but otherwise exactly like them. It has a north-seeking and a south-seeking pole like any other magnet, but from the laws of attraction and repulsion we see that, curiously enough, its south-seeking pole must be at Peary's end, and its north-seeking pole at Amundsen's end. These magnetic poles are not exactly at the geographical poles. One of them is in North America near Hudson's Bay and the other is nearly opposite.

Since the lines of equal declination and of equal dip are not true circles, the magnetization of the earth must be somewhat irregular. Furthermore, the positions of its magnetic poles are known to be changing slowly from year to year. Why these things are so, and, for that matter, why the earth is magnetized at all, is not yet known.

QUESTIONS

1. Does a magnet ever have more than two poles?
2. In what direction did Peary's compass point when he reached the North pole?
3. How far is the magnetic pole from the geographical North pole?
4. How can you tell whether or not a steel rod is a permanent magnet?
5. Why are knives, files, and scissors sometimes found to be magnetized?

6. Will a magnet attract a tin can? Explain.
7. Would a magnet floating on a cork in a dish of water float toward the north, as well as turn north and south?
8. What advantage is there in making a magnet in the shape of a horseshoe?

237. The field around a magnet. Michael Faraday (1791–1867) was the first to see that a true understanding of the action of magnets could be had only by studying the empty space around them, as well as the magnets themselves.

One way to do this is to lay a stiff piece of paper over a magnet and sprinkle iron filings on it (Fig. 200). When the paper is tapped lightly so as to shake the filings about a little, they arrange themselves in regular lines leading from one pole to the other. This is because each

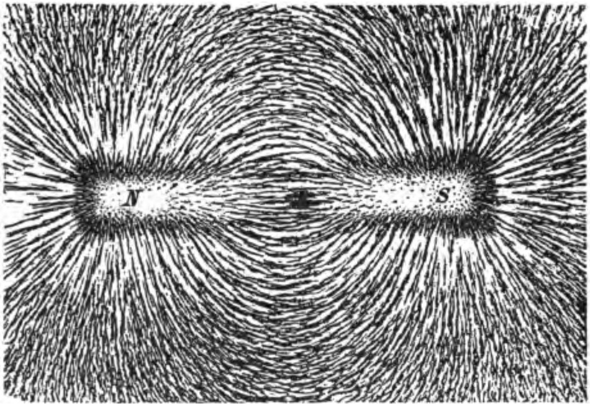


FIG. 200. — Magnetic lines of force around a bar magnet.

filing gets slightly magnetized by the influence of the original magnet, and sets itself in the direction in which a tiny compass needle would lie if it were at the same place. This can be verified by actually using a small compass instead of the filings. The lines can be mapped in this way, but it is not as quickly done.

In this way, Faraday drew what he called **lines of force** around a magnet. A line of force may be defined as a line

which indicates at its every point the direction in which a north-seeking pole is urged by the attractions and repulsions of all the poles in the neighborhood. When lines of force are thought of in this way, they should have little arrowheads on them, pointing in the direction of their journey from a north-seeking pole to a south-seeking pole. We shall find this conception of lines of magnetic force or **magnetic flux** a convenient way of remembering how a magnet will affect other magnets in its vicinity.

238. Lines of force like elastic fibers. Faraday himself thought of these lines of force as having a much more real meaning than this. He thought of them as actually existing throughout the space around every magnet, even when there are no filings to show them. He believed that they represent a real state of strain in the **ether** (see section 187), in which all material bodies are immersed. Even now we know very little about what the ether really is. We know simply that it is not a kind of matter, but something much more subtle and fundamental.

At any rate, these lines of force of Faraday's *act as if they were stretched fibers in the ether* which are continually trying to contract and are thus pulling on the poles at their ends. They also *act as if they were trying to swell up sidewise as they contract*, and thus seem to crowd each other apart. It is not easy to see *why* lines of force have these properties, but once the properties are assumed (as rules of the game), it is easy to reason out from them *what will happen* in many practical cases.

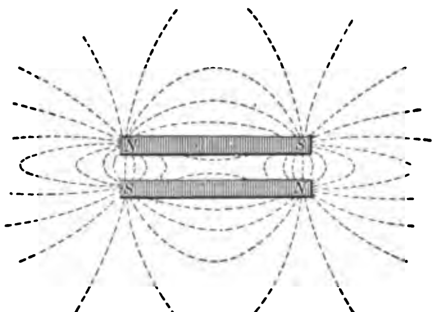


FIG. 201. — Lines of force between two unlike poles.

For example, if two magnets are placed with their *unlike* poles together and their lines of force traced with iron filings, the result will be as shown in figure 201. If we assume that the lines of force tend to contract, it is easy to see that two unlike poles must *attract* each other. *Like* poles, however, would show a field of force as shown in figure 202, and if we assume that lines of force squeeze against each other sidewise, and tend to separate, evidently two like poles must *repel* each other. This is not an explanation of *why* these things happen; it is, however, an easy way of seeing *what* will happen, and it will be useful later on.

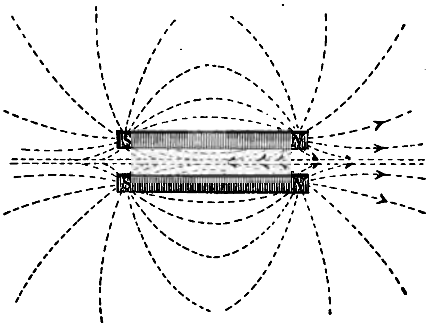


FIG. 202.—Lines of force between two like poles.

239. Induced magnetism. If we plunge one end of a piece of unmagnetized soft iron into some iron filings, it does not attract them, but if we bring near it a permanent magnet, as shown in figure 203, the soft iron becomes a magnet and attracts the filings. When the permanent magnet is removed, the soft iron loses its magnetism, and drops the filings.

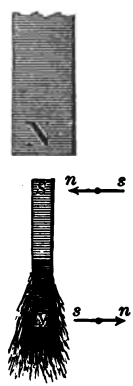


FIG. 203. — A magnet by induction.

A piece of iron which is magnetized by being near a magnet is said to be **magnetized by induction.** If the pole of the magnet, which was brought near the iron, was a north-seeking pole, the induced magnet can be shown by a compass to have a *N*-pole away from the magnet and a *S*-pole near the magnet.

Experiments show that very soft iron quickly becomes magnetized by induction and quickly loses its magnetism when removed from the field. Hardened steel, however, is magnetized with difficulty, but retains its magnetism well. For

this reason the magnets used in telephones and magnetos are made of hardened steel.

240. Permeability. Although a magnet will act through a vacuum or through glass or wood, yet the magnetic flux seems to prefer soft iron to any other medium.

We can show this by the following experiment. We will take a horseshoe magnet and lay across its poles a sheet of stiff paper, and then bring up a mass of iron filings under the paper. The iron filings will cling under the poles as shown in figure 204. If we slip a plate of glass between the paper and the poles at AA, most of the filings still stick, but when we substitute an *iron plate* for the glass, most of the filings drop off immediately. This shows that an iron plate screens the region beyond from the magnetic action.

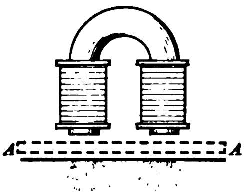


FIG. 204. — Iron is more permeable than glass.

Lord Kelvin has called the ease with which lines of force may be established in any medium as compared with a vacuum, the **permeability** of the medium. Thus iron has a permeability several hundred times greater than air. When a watch is brought near a powerful magnet, its balance wheel is often magnetized. This disturbs its working. To protect it from such magnetic disturbances a good watch is often inclosed in a soft iron case.

241. Theory of magnetism. Our present theory of magnetism was suggested by the following experiment.

Let us harden a knitting needle or a piece of watch spring by first heating it red hot, and then plunging it into cold water. Then let us magnetize it and mark the N-pole. If we now break it near the middle

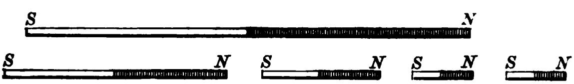


FIG. 205. — A broken magnet shows poles at the break.

where it does not show any magnetism, we shall find, by bringing the broken ends near a compass needle, that we have an N-pole and an S-pole as indicated in figure 205. If we repeat the process, we shall

find that each time the magnet is broken, new poles are formed at the break.

A magnet can be broken into a great number of little magnets. A glass tube full of iron filings can be magnetized, but when shaken, it loses its magnetism. Any magnet loses a part or all of its power if it is heated red hot, jarred, hammered, or twisted.

All these facts point to a **molecular theory of magnetism,** which was suggested by a Frenchman, Ampere, and elaborated by a German, Weber, and an Englishman, Ewing. Every molecule of a bar of iron is supposed to be itself a tiny permanent magnet — why, no one yet knows. Ordinarily, these molecular magnets are turned helter-skelter throughout the bar (Fig. 206), and have no cumulative effect that can be noticed outside the bar. When the bar is magnetized, however, they get lined up more or less parallel (Fig. 207), like soldiers, all facing the same way. Near the middle of the bar the front ends of one row are neutralized by the back ends of the row in front; but at the ends of the bar a lot of unneutralized poles are exposed, north-seeking at one end and south-seeking at the other. These free poles make up the active spots which we have called the poles of the magnet.

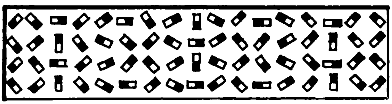


FIG. 206. — Unmagnetized bar.

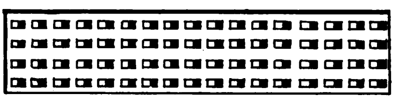


FIG. 207. — Magnetized bar.

On this theory it is easy to see that when a magnet is broken in two without disturbing the alignment of the molecular magnets, the new poles which appear at the break are simply collections of molecular poles that have been there all the time, but are now for the first time in an independent, recognizable position.

It will also be evident that, if this theory is true, there

is a perfectly definite limit to the amount of magnetism a given piece of iron can have. For when all the molecular magnets are lined up in perfect order, there is nothing more that can be done, no matter how strong the magnetizing force may be. Such a magnet is said to be **saturated**.

SUMMARY OF PRINCIPLES IN CHAPTER XIII

Like poles repel each other.

Unlike poles attract each other.

The earth is a magnet with its "south-seeking" pole at Peary's end.

Lines of force tend to contract and swell sidewise; that is, there is tension along them, and compression at right angles to them.

QUESTIONS

1. If two bar magnets are to be kept side by side in a box, how should they be arranged? Why?
2. If a magnetic needle is attracted by a certain body, does that prove that the body is a permanent magnet?
3. What is meant by the "aging" of magnets?
4. How must a ship's compass box be supported so as to remain steady during the rolling of the ship?
5. A long soft iron bar is standing upright. Why does its lower end repel the north pole of a compass needle?
6. Does hammering the bar while it is in the position described in problem 5 increase or decrease the effect? Why?
7. Why are the hulls of most iron ships permanently magnetized? What determines the direction in which they are magnetized?
8. How can the compass on an iron ship be "compensated" for the induced magnetism in the ship?
9. The Carnegie Institute has a special ship built almost without iron. What kind of a survey of the world do you suppose it is made for? What is the advantage of such a ship for this purpose?
10. How does a jeweler demagnetize a watch?
11. What effect does the angle of dip have on the horizontal intensity of the earth's magnetism at any point?

CHAPTER XIV

THE BEGINNINGS OF ELECTRICITY

Frictional electricity — conductors and insulators — positive and negative charges — electroscope — frictional electric machine — the lightning rod — induction — Leyden jar — electrophorus — theories as to nature of electricity.

242. Electricity by friction. As far back as 600 B.C., Thales of Miletus, one of the "seven wise men," knew that the yellow resinous substance called amber, of which pipe-stems and jewelry are now often made, would, when rubbed, attract bits of paper or other light objects. We now know that many other substances, such as rubber, glass, and sulphur, have the same property. Any one can observe this on a cold, dry morning after combing his hair vigorously with a hard rubber comb. The comb will then support long chains of bits of paper. Another way to show this is to scuff one's feet on a carpet, or to rub a cat's back. In either case, if the knuckle is brought near a gas fixture, tiny sparks will pass. Since amber, in common with gold and certain bright alloys, was called "electron," by the Greeks, these phenomena were many years later named by Gilbert "electric," that is, "amberous," phenomena.

243. Electric *vs.* magnetic attraction. These electric attractions are in many ways so much like magnetic attractions that it was not until the sixteenth century that it was clearly seen that two very different kinds of phenomena are involved. *Magnetization* can be produced only in three metals, iron, nickel, and cobalt, and in one or two uncommon alloys, while *electrification* can be produced by rubbing almost any sub-

stance, especially non-metals. A magnetized body always has at least two poles where its magnetism is more or less concentrated, and these poles are *unlike*, for if one of them attracts the north-seeking end of a compass, the other will always repel it. A metallic body electrified by friction will ordinarily not have its properties concentrated in spots, and all parts of it will act very much *alike* in their attracting power. Nevertheless, we shall presently see that there are two kinds of electricity, just as there are two kinds of magnetic poles.

244. Conductors and insulators. Some substances will conduct electricity, while others will not. Thus a metal sphere can be charged with electricity by touching it with some electrified substance, such as a stick of sealing wax which has been rubbed with a cat's skin, *if* the sphere is suspended by a dry silk thread, but not if suspended by a wire. In the latter case just as much electricity gets into the sphere as in the former, but it all runs out again through the wire. So we distinguish between **conductors**, the best of which are the metals, and **non-conductors** or **insulators**, such as dry silk, glass, hard rubber, sulphur, porcelain, paraffin, and resin. It is to prevent the leakage of the electricity in the conductor that electric light, telephone, and telegraph wires are supported on glass or porcelain knobs called "insulators."

There is no sharp line between conductors and insulators; most substances conduct a little, and even the good conductors vary greatly in conductivity.

In the following table a few common substances are arranged according to their insulating powers.

INSULATORS	POOR CONDUCTORS	GOOD CONDUCTORS
Amber	Dry wood	Metals
Sulphur	Paper	Gas carbon
Glass	Alcohol	Graphite
Hard rubber	Kerosene	Water solutions of
Dry air	Pure water	salts and acids

It will be noticed that the substances which can be easily electrified by friction are all insulators. One reason for this is that when electricity is generated at any point on a body by rubbing, it stays there and makes its presence known, if the body is an insulator; but if it were a conductor, the electricity would leak away at once.

It will also be noticed that those substances which are good conductors of electricity are also good conductors of heat. This curious fact, long not understood, seems to be due to the fact that both heat and electricity are carried through metals by a swarm of tiny particles, called **electrons**, which drift about between the much larger molecules of metal like wind through a forest.

245. Positive and negative electricity. If we hang up in a stirrup, suspended by a silk thread, a glass rod which has been rubbed with silk, and then bring near one end of it another glass rod which has also been rubbed, they repel each other (Fig. 208). In a similar way two hard rubber rods or sticks of sealing wax repel each other. But when we bring a rubbed stick of sealing wax near a rubbed glass rod in the stirrup, they attract each other.

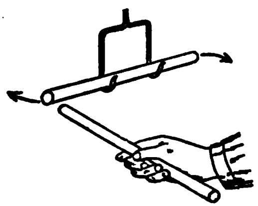


FIG. 208. — Two electrified rods repel each other.

From such experiments as these we have come to distinguish between two states of electrification. We call one kind "vitreous" (glass) electricity or **positive** electricity, and the other "resinous" electricity or **negative** electricity. Bodies charged with the same kind of electricity repel each other, and bodies charged with different kinds of electricity attract each other. That is,

Like charges repel and unlike charges attract.

246. How to detect electricity. To test the electrical condition of a body we use an **electroscope**. A simple form of

electroscope consists of a pith ball hung by a silk thread from a glass support (Fig. 209).

If an uncharged body is brought near the pith ball, nothing happens. If a positively charged body is brought near the pith ball, the latter is attracted, becomes itself positively charged, and is then repelled. Then if a negatively charged body is brought near, the positively charged pith ball is attracted, but when it touches, it becomes negatively charged and flies back. If we now bring a negatively charged body near, the negatively charged pith ball is repelled. If, then, we know what the nature of the charge on the pith ball is, and find that a body repels it, we know the body must be charged the same way. If there is attraction, we cannot be sure whether the body is uncharged or oppositely charged.

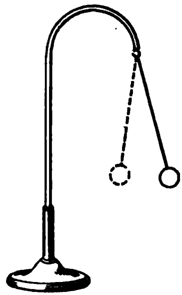


FIG. 209.—Pith-ball electroscope.

A more reliable form of electroscope is the so-called "gold-leaf" **electroscope**, although nowadays they are quite commonly made of two aluminum leaves hung from a brass rod. These are usually mounted in some sort of a glass case, as shown in figure 210.

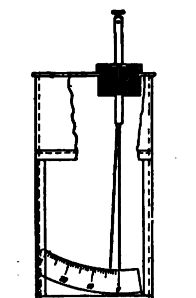


FIG. 210.—The aluminum-leaf electroscope.

When one brings near the top of the brass rod a charged glass rod, the aluminum leaves separate and hang like an inverted V. If the rod is removed, the leaves come together again. If, however, one actually touches the charged rod to the electroscope, the leaves separate and stay apart.

The electroscope is then said to be charged. If we bring near a positively charged electroscope a positively charged body, the leaves will fly farther apart; but if the body brought near has a negative charge, the leaves will fall toward each other. In either case they will return to their original charged position when the outside charged body is taken away. So with an electroscope one can tell the electrical condition of a body.

With such an electroscope it is possible to learn much about electrified bodies. For example, when an insulated conductor is rubbed, it becomes charged with electricity; so we conclude that all bodies become electrified by friction. If we stand on an insulated stool, while we rub a glass rod, our body becomes negatively charged; and by rubbing sealing wax with cat's fur, we become positively charged. In general, ***whenever two different substances are rubbed on one another, one becomes positively charged with electricity, while at the same time the other is negatively charged.***

247. Frictional electric machine. All the early forms of electrical machines were frictional, and such machines are still used for demonstration purposes. A circular glass plate is mounted firmly on an axle, so that it can be turned between two silk-covered cushions, which are pressed against the glass by springs. The charge on the glass is drawn off by a metal comb which is supported on a glass rod. When the plate is rotated, it becomes positively charged and this charges the metal comb positively; at the same time the rubbers become negatively charged and should be connected with the ground by a wire or chain, that is, **grounded**, so that the negative charges can escape.

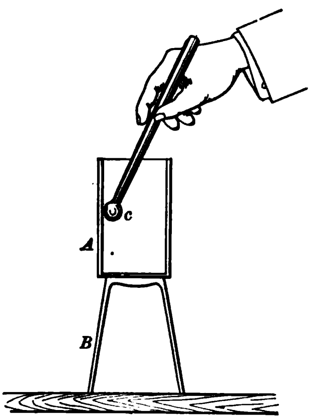


FIG. 211. — Charging a metal can.

248. Distribution of electricity on a conductor. Let us place a metal can, such as is used for heat measurements, on a glass plate as shown in figure 211, and connect it with an electrical machine by a wire. After we have charged the can as much as possible, we may test it at various points by means of a little metal disk or ball mounted on an insulating handle and known as a proof plane. If we touch it to the outside surface of the charged can and bring it near the knob of

a charged electroscope, and then repeat the test, touching it to the inside surface, we find that there is a strong charge on the outside but none on the inside.

Such experiments show that the charge is entirely on the outer surface of a conductor and that its greatest density is at the corners and projecting points. In fact the density of the charge at sharp points is so great that the charge will escape into the air easily at such points.

If we attach a tassel of tissue paper to the insulated conductor of the electric machine and charge it, the little paper streamers repel each other and stand out in all directions, but if a needle point is held near them, they fall together at once.

We may fasten to a friction machine an "electric whirl," balanced on a pin point. When the machine is started, the whirl turns as shown in figure 212.

If a bent point is attached to the machine, and a candle flame is held near the point, the so-called "electric wind" may blow the candle flame aside. It is not, however, the electricity itself that blows the candle, but the surrounding air which is in some way set in motion by the discharge.

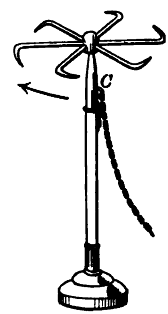


FIG. 212. — Electric whirl.

Such experiments show that a conductor can be charged or discharged more easily at a sharp point than at a rounded surface.

249. Lightning and lightning rods. For a long time people supposed that thunder and lightning were caused by the combustion of some kind of gas in the clouds. But when electricity began to be studied, it occurred to some philosophers that lightning might be an electrical phenomenon. Thus we find Benjamin Franklin in his notebook, under the date of November 7, 1749, making a list of the respects in which lightning resembled electric sparks, such as "giving light, color of the light, crooked direction, swift motion, being conducted by metals, crack or noise in exploding, rending bodies it passes through, destroying animals, heating metals and kindling inflammable substances, and its sulphurous smell"

(now known to be due to ozone). He then wondered if lightning, like electricity, could be drawn off by points. "Since they agree in all the particulars wherein we can already compare them, is it not probable that they agree likewise in this? Let the experiment be made."

But this was not easy. Franklin thought he would require a tower or steeple high enough to reach into the clouds themselves, and his friends set about raising the money to build one by giving popular lectures on electricity all over the country. In the meantime two Frenchmen were bold enough to try the experiment with insulated pointed rods less than a hundred feet high, and were successful in drawing off sparks from the lower ends of the rods during thunder showers. But Franklin was not satisfied, because the rods did not reach into the thunder clouds, and might have been electrified some other way. Suddenly in 1752 a new idea flashed into his mind, and he set about making his famous kite. The result is known to every one. Almost the most wonderful part of it was that Franklin was not killed at once. Within a year one Richman was killed while making a similar experiment in St. Petersburg.

So Franklin invented the lightning rod to conduct electricity safely from the clouds to the earth. Nowadays in cities where the houses are built in blocks with frameworks, tops, and cornices of metal, the lightning rod is not much used. But tall chimneys, church steeples, and isolated houses are often provided with lightning rods.

It should be remembered that unless a lightning rod is put up with considerable care, it is a menace rather than a protection. In particular, its lower end must be well "grounded," as by soldering to large copper plates buried in *damp* soil, and no part of the rod should turn a sharp corner. If these precautions are not observed, a lightning rod will often discharge into the house itself, rather than into the ground, the electricity which it has attracted.

QUESTIONS

1. Compare the behavior of a magnetic pole with the behavior of an electrically charged body.
2. Does a freely swinging charged body take a definite direction?
3. What becomes of the mechanical energy exerted in rubbing a glass rod to electrify it?
4. What kind of electricity is generated by rubbing a fountain pen on woolen cloth?
5. Why do experiments with frictional electricity work better on a cold, dry winter day?
6. Does one remove magnetism from a magnet by touching it with iron?
7. Faraday built a large box and lined it with tin foil. He then took his most sensitive electroscope into the box and found that even when the outside of the tin foil was so charged that it sent forth long sparks, he could not observe any electrical effects inside. Explain.
8. What evidence have you that the human body is a good conductor?

250. Charging by induction. If one brings a positively electrified ball near an insulated conductor, such as a metal cylinder on a glass support, and then removes it again, the cylinder is not electrified. But if, while the electrified body is near, one touches the cylinder with his finger or a grounded wire for an instant, the cylinder is found to be negatively charged after the charged body has been removed. If one repeats this experiment using a negatively charged ball, the metal cylinder becomes positively charged. Since the electricity is not diminished in the ball, we must look to the cylinder for the electricity. Charges produced in a conductor by virtue of its proximity to a charged body are called **induced charges.**

This process of charging by induction may be explained as follows. When the positively charged ball is brought near the cylinder, the positive and negative electricity in the cylinder are distributed as shown in figure 213.

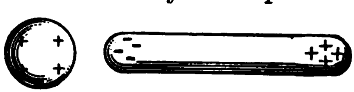


FIG. 213. — Charging by induction.

When one touches the cylinder, the positive electricity,

which is repelled, finds its way to the ground through the body, but the negative electricity remains **bound** (Fig. 214). It does not flow off when the conductor is touched, but is held by the presence of the charged body.

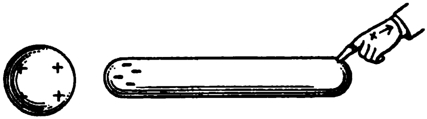


FIG. 214. — Bound charge.

This helps us to understand the gold-leaf electroscope. When a charged body is brought near the knob of the electroscope, the leaves separate because they are charged by induction with the same kind of electricity as the charged body (Fig. 215). If the electroscope is charged by contact positively and a positively charged body is brought near, it repels more of the positive electricity into the leaves and so they diverge more widely. On the other hand if a negatively charged body is brought near, it draws some of the positive electricity up into the knob, and the leaves come together more or less according to the amount of the charge.

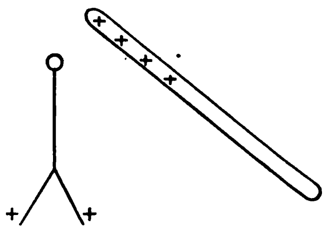


FIG. 215. — Charging an electroscope by induction.

251. Condenser. In many practical applications of electricity, it is necessary to increase the capacity of a conductor for holding electricity. This is done in what is called a condenser.

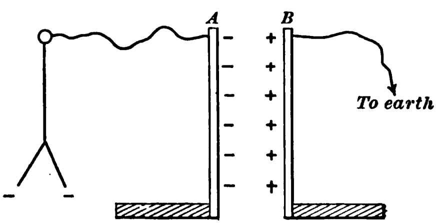


FIG. 216. — Action of a condenser.

Let us arrange a metal plate on an insulating base and connect the plate by a wire to an electroscope, as shown in figure 216. If we charge the plate *A*, we see the leaves of the electroscope diverge. We will now

bring up a second metal plate B similar to plate A, but connected with the ground. As we bring the plate B near plate A, the electroscope leaves begin to fall together, but if we remove plate B again, the leaves separate as before.

Let us now bring the plate B back to a position near plate A, and charge plate A until it shows the same deflection as before. It will be evident that the capacity of plate A for holding electricity is much increased by being close to a similar grounded plate B.

We may also show the influence of the insulating material between the conducting plates, by introducing a pane of glass. The leaves of the electroscope fall nearer together, but rise again when the glass is removed. This shows that the capacity of the condenser is increased by the glass plate.

A combination of conducting plates separated by an insulator is called a **condenser**. The capacity of a condenser for holding electricity is proportional to the size of the plates and increases as the distance between them decreases. It also depends on the nature of the insulator, or **dielectric**, as it is called. Mica and paraffin paper are much used in commercial work.

252. Leyden jar. At the University of Leyden in Holland, as early as 1745, they used a condenser in the form of a wide jar or bottle (Fig. 217), coated inside and out with tin foil. Inside the jar, and connected at the bottom to the inside coating, is a rod with a knob on top. If one allows a charge of positive electricity to jump to the knob the positive electricity on the inner lining attracts through the glass the negative electricity of the outer coating, while at the same time the compensating positive electricity originally in the outer coating is repelled and escapes through its support or the hand which holds it. It is possible to make a great number of sparks jump to the knob before it ceases to receive them. Then the jar is **charged**. If one connects the outer coating and the knob by a metal wire, the elec

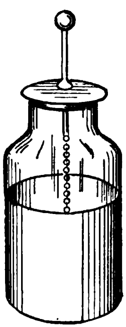


FIG. 217. — Leyden jar.

trical strain or pressure is released with a bright crackling spark. If the jar is discharged through a piece of paper, the spark makes a hole in the paper. If one makes the connection through his own body, he feels a lively sensation, known as a shock.

253. Hydraulic analogy of a condenser. We may illustrate a condenser by two standpipes filled to different levels with water, as shown in figure 218. The coatings of the condenser correspond to the standpipes. The pipe *A*, with the water standing at a higher level, represents the positively charged plate or coating, while the other pipe *B* is the negatively charged plate. The connecting pipe at the bottom of the tanks corresponds to the wire connecting the coatings. When the connection is made, the water rushes through the pipe and equalizes its levels very quickly. This represents the discharge of the condenser.

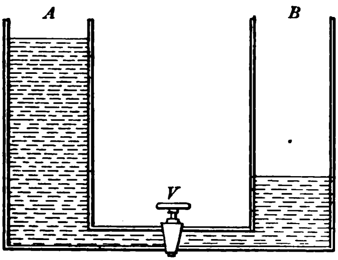


FIG. 218. — Hydraulic analogy of a condenser.

When the valve *V* in the pipe is first opened, the water rushes through so fast that it usually overdoes things, and rises to a higher level in *B* than in *A*. Then it flows back again and so on, oscillating back and forth until the motion dies out because of friction in the pipe. In much the same way, when a condenser is short-circuited, the discharge of electricity goes too far and charges up the condenser the other way. Then it discharges back again, and so the electric charges oscillate very quickly back and forth until the motion of the electricity dies out because of something akin to friction, called the **electrical resistance** of the wire. The technical way of describing this is to say that the discharge of a condenser is **oscillatory**.

254. Induction machines for producing electricity. The simplest machine for producing electricity by induction is the **electrophorus**. It consists of a hard rubber disk and another somewhat smaller metal disk, which is provided with an insulating handle (Fig. 219).

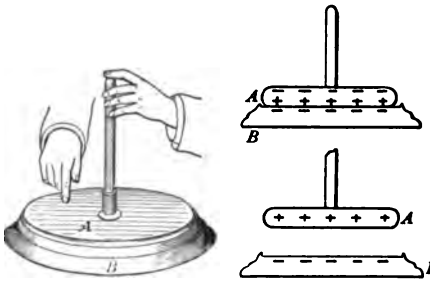


FIG. 219. — Electrophorus.

If we rub the hard rubber plate of an electrophorus with cat's fur, we find it is charged negatively. Then we place the metal disk on the plate and touch the finger to the metal disk so as to "ground" it. When we lift up the disk and bring it near the knuckle, or the knob of a Leyden jar, a spark jumps across the gap. We may charge a Leyden jar with an electrophorus by repeating this process again and again.

When the rubber plate is electrified, it becomes negatively charged. When the metal disk is placed upon it, a positive charge is attracted to the lower surface of the disk next to the plate, while the negative electricity is repelled. When we touch the metal disk, this negative electricity escapes through the hand to the ground. In this process the disk becomes charged positively throughout. After the rubber plate is once charged, any number of charges can be obtained from the electrophorus, without producing any appreciable change in the charge on the plate. This is because the energy comes from the agent who lifts the disk.

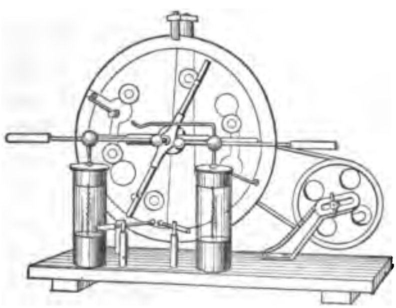


FIG. 220. — Toepler-Holtz machine.

255. Toepler-Holtz machine. Among the ma-

chines which make use of this principle of induction to produce electricity is the so-called Toepler-Holtz machine, shown in figure 220. The details of construction are so many, and the explanation of its operation is so complex, that it is left to the special books on electricity. If one of these machines is in working order, many entertaining experiments can be done with it.

256. Theories as to the nature of electricity. To explain these electrical phenomena, du Fay, a Frenchman, assumed that there were in all bodies two fluids, namely vitreous electricity and resinous electricity. When these are present in equal quantities, they neutralize each other. If a glass rod is rubbed by silk, the silk, which has a greater "affinity" for the resinous fluid than the glass, absorbs some of it from the glass, and at the same time the glass, having a greater affinity for the vitreous fluid than the silk, absorbs some from the silk. So each body gets an excess of its preferred fluid and becomes charged.

Later Franklin suggested that there was only one kind of fluid, namely, vitreous electricity, of which a certain amount "belonged" in every body. If it had an excess, it was what had been called vitreously charged. If it had less than enough, it was resinously charged. This led to the terms "positive" and "negative" charge, which are still in use.

Lately, we have come back to something nearer du Fay's idea. We do not think of electricity as a kind of matter, as the word "fluid" indicates, but we believe that there are two kinds, a negative or resinous kind occurring in very small lumps which we now call **corpuscles** or **electrons**, and a positive kind of a different nature, not yet understood. Even in the modern electron theory, however, there are some who prefer to believe with Franklin that there is only one kind of electricity, namely electrons, which may be present either in excess or in defect. If this turns out to be true,

Franklin's only mistake was that he hit on the wrong kind of electricity as positive.

It makes very little difference whether we talk and think in terms of the one-fluid or the two-fluid theory, inasmuch as everything we know can be expressed either way, and we do not yet know which is right.

257. Conclusion. Practically all that people knew about electricity up to the beginning of the nineteenth century has been briefly outlined in this chapter in very much the order in which it was discovered. Few discoveries were made, and they dealt only with electricity at rest (**electrostatics**). Almost the only useful electrical invention was the lightning rod, and its usefulness has been much overestimated. The most useful instrument which had been devised was the condenser. Nevertheless, the people of the eighteenth century were fascinated by electricity. It was the most exciting topic with which scientific men dealt; it was lectured about and shown off to large audiences, and was as much talked about by everybody as radium or wireless telegraphy have been recently. But it was merely a plaything in laboratories.

In the last half of the nineteenth century, as we shall see in the following chapters, electricity suddenly leaped into a commanding position in the arts and engineering. Probably no more spectacular service has ever been rendered to the welfare of mankind by what practical men like to call "pure science." The story of this development is a most convincing answer to those who, even now, distrust "pure science" as "impractical" and "useless."

SUMMARY OF PRINCIPLES IN CHAPTER XIV

All bodies can be electrified by *friction*, becoming charged either positively (vitreously) or negatively (resinously).

Like charges repel each other.

Unlike charges attract each other.

All conductors can be electrified by *induction*, showing both a positive and a negative charge in different places. Of these one is *bound* by the inducing charge, but the other is *free*.

QUESTIONS

1. Why cannot a Leyden jar be appreciably charged if the jar stands on a glass plate?
2. If a charged Leyden jar is placed on a glass plate, why does one not get a shock if he touches the knob?
3. How would you arrange four Leyden jars to get increased capacity?
4. How would you arrange four Leyden jars to get as long a spark as possible?
5. If an insulated metal globe is negatively charged, how can any number of other insulated globes be positively charged?
6. If an insulated metal globe is negatively charged, how can any number of other insulated metal globes be negatively charged?
7. In the experiment shown in figure 214, why must the finger be removed before the removal of the charged body?

CHAPTER XV

BATTERY CURRENTS

The voltaic cell — action in a cell — hydraulic analogy — defects of simple cell — commercial cells.

Magnetic field around a current, and around a coil — electromagnet — electric bell — telegraph.

BATTERIES

258. Beginnings of the electric battery. For nearly two thousand years friction and induction were the only methods known for producing electricity. But, in 1786, an unexpected observation of an Italian anatomist, Galvani, in Bologna, started a series of most important discoveries and inventions. He observed that the legs of frogs which he had been dissecting twitched every time there was a discharge from his electric machine. Later he found that if strips of two different metals, such as copper and zinc, were fastened together like an inverted V, and their free ends applied to frogs' legs, there were the same nervous twitchings as followed the discharge of electricity. Therefore he concluded that he had found a new way of producing electricity. He thought the electricity was formed at the contact of the dissimilar metals.

While investigating this question, Volta invented a chemical method of producing electricity continuously, called an **electric battery**.

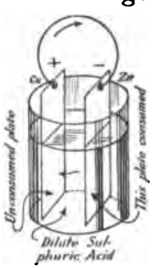


FIG. 221. — Voltaic cell.

259. Voltaic battery. A glass tumbler, with a strip of zinc and a strip of copper dipping into dilute sulphuric acid (Fig. 221), is one form of **voltaic cell**, and when several cells are combined, they constitute a **battery**.

To show that the copper and zinc strips are each charged with electricity, we will connect six such cells in series as shown in figure 222. To detect the feeble charge we will put a 3-inch disk on the top of the brass rod of the aluminum-leaf electroscope. Then we will take another similar disk which is provided with an insulating handle and has a thin coating of shellac on the bottom, and place this disk on top of the other. This forms a **condensing electroscope.** If we touch the wires leading from the zinc and copper strips of the battery to the lower and upper disks of the condenser, as shown in figure 222, and then remove the wires and lift off the upper disk, we find that the leaves of the electroscope diverge. If we bring a charged stick of sealing wax near the electroscope, the leaves spread still farther apart, which shows that the electroscope and the **zinc are negatively charged.**

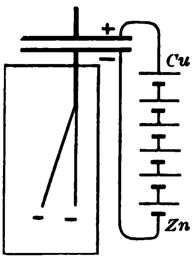


FIG. 222. — To show charges on plates of voltaic cell.

If we repeat the experiment with the wires reversed, we can show that the **copper** is **positively charged.**

The copper (or the carbon which often replaces it) is called the positive electrode or **+ pole,** while the zinc is called the negative electrode or **− pole.** The solution in the cell is called the **electrolyte.** When the poles of a cell are joined by a conductor, we have an electric path or **circuit** consisting of the electrodes, the electrolyte, and the metallic conductor joining the poles. If a bell or lamp is to be operated by an electric battery, it is so connected that the electricity passes through it as a part of the circuit. When this circuit is broken at any point by a switch, key, or push button, so that no electricity jumps the gap, the circuit is said to be **open.** When the switch or key is closed so as to make a continuous path, the circuit is said to be **closed** or **made.**

260. Action of an electric cell. We have already seen that when a Leyden jar is discharged, or any two charged bodies are connected by a wire, there is what we call a flow of electricity; that is, an **electric current.** By convention we say that the *electricity in the connecting wire flows from the positive to the negative conductor.* In a single electric cell we shall speak,

therefore, of the electricity as flowing through the *outside* circuit *from* the copper or carbon electrode (+ pole) *to* the zinc electrode (− pole). *Inside* the cell the electricity must evidently flow "uphill" through the solution or electrolyte back to the copper electrode. We shall see presently why it is able to flow uphill inside the cell.

To understand a little better just what is happening inside the cell, let us dip a strip of ordinary zinc into very dilute sulphuric acid. We shall see bubbles rising from the zinc and coming to the surface of the acid. These bubbles are a gas called **hydrogen.** If we leave the zinc in the acid, it gradually dissolves, leaving behind only a few insoluble impurities.

If we repeat the experiment, using a copper strip, we shall find no action; but if we put both the zinc and the copper strips into the acid and connect them with copper wires to some instrument that indicates a current of electricity (a galvanometer), we see that a current is produced, and that bubbles are coming from both the copper and the zinc strips.

Next we will remove the zinc strip and rub a little mercury on it. The mercury clings to the zinc and can be spread over its surface. Such a union of a metal with mercury is called **amalgamation.** If this amalgamated zinc is used in the cell, no bubbles are formed on it. When the circuit is closed, bubbles rise from the copper plate, and when the circuit is broken or open, these bubbles stop. A galvanometer in the circuit shows a current as before, but now the amalgamated zinc is consumed only when the circuit is closed. The copper is not consumed by the acid at all.

In general it can be said that the electric current depends on the difference in the chemical action of the acid on the two metals used as electrodes. The metal which is dissolved or acted upon by the acid is the negative electrode; the metal which is apparently unchanged and from which the hydrogen bubbles rise while the circuit is closed is the positive electrode.

261. The chemistry of the cell. In chemistry we learn that sulphuric acid is made up of two parts hydrogen, one part sulphur, and four parts oxygen, as expressed by the symbol H_2SO_4. When sulphuric acid is dissolved in water, some of it breaks up into two parts, $\overset{++}{H_2}$ and $\overset{--}{SO_4}$. These

two parts, called **ions**, carry opposite kinds of electricity. The $\overset{++}{H_2}$ is positively charged and the $\overset{--}{SO_4}$ is negatively charged.

When zinc (Zn) is placed in the acid, a little of it dissolves, becoming zinc ions ($\overset{++}{Zn}$), which unite with the $\overset{--}{SO_4}$ ions to form zinc sulphate ($ZnSO_4$). The displaced hydrogen ($\overset{++}{H_2}$) goes to the copper plate, gives up its charge to the plate, and then rises as bubbles of gas. It is important to remember that the *positively* charged part of the electrolyte ($\overset{++}{H_2}$) *goes with the current* through the cell. The electric current will flow through the wire from the copper to the zinc as long as the chemical action is maintained. Thus we see that it is the energy of the chemical action which forces the electricity to run uphill inside the cell. In this way chemical energy is transformed into electrical energy.

In a good commercial cell the chemical action takes place only when the cell is delivering electrical energy. The rate at which this energy is delivered by the cell determines the rate at which the zinc is used up; just as the rate at which steam energy is delivered by a boiler determines the rate of coal consumption. *Zinc is, then, the fuel of the electric cell.*

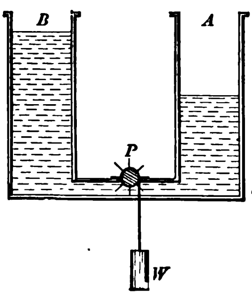


FIG. 223.—Water at different levels.

262. Electric currents and water currents. Although it must not be supposed that electricity *is* a material flowing through the circuit as water flows through a pipe, yet it will greatly help us to form a mental picture of the situation if we compare electric currents with water currents. In figure 223 we have two tall open vessels containing water. These are connected by a pipe which contains a pump driven by the weight *W*. The water will evidently be pumped from *A* to *B*.

until the back pressure on the pump due to the higher level of the water in *B* is enough to balance the weight *W*. This difference in level does not depend on the size of the vessels.

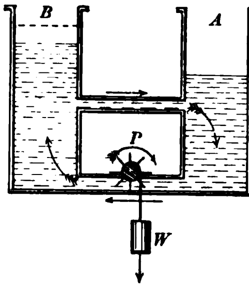


FIG. 224. — Water in circulation.

Suppose now that the vessels are connected by a second pipe, as shown in figure 224. Then the difference in levels will cause the water to flow from *B* to *A*. The water level in *B* drops a little and that in *A* rises, so that the difference in levels between *A* and *B* becomes less. When the back pressure against the wheel of the pump is thus reduced, the weight drops and drives the water around the circuit. This will continue as long as the weight can move downward.

The difference in level in *A* and *B*, in figure 223, represents the difference in the electrical condition of the two electrodes, copper and zinc. This is called the **difference of potential** between the positive and negative poles of the cell. The pump represents the chemical action of the acid on the zinc, which produces this difference of potential. Figure 223 is, then, analogous to the cell with its circuit open.

The cell with its circuit closed is represented by figure 224. The tube connecting *A* and *B* represents the outside circuit between the copper and the zinc. The circulation of the water represents the flow of electricity. The rate at which the water circulates depends on the difference in level which the pump can maintain; that is, on the power of the pump. Similarly the rate of flow of electricity depends on the **electromotive force** which the chemical action of the acid and zinc can maintain. Furthermore, the rate of flow of the water depends on the friction in the connecting pipes, and similarly, the rate of flow of the electricity depends on the electrical friction or **resistance** of the circuit. Finally, just

as the energy needed to circulate the water comes from the action of gravity on the weight, so *the energy needed to drive the electric current is supplied by the chemical changes which take place at the electrodes.*

263. Two defects in a simple cell. Volta's simple cell, which has been described, was soon found to have two defects, local action and polarization. When ordinary zinc is used, bubbles of hydrogen are formed at the surface of the zinc strip even before it is connected with the copper. This means a wearing away of the zinc to no purpose, and is called **local action.** It is due to impurities, such as iron or carbon, embedded in the zinc. These impurities form with the zinc a minute voltaic cell, as shown in figure 225. The local current flows from the iron or carbon directly to the zinc and then back through the acid to the iron again. In this process, the zinc is eaten away near the impurity, and hydrogen is set free. To avoid this useless wasting away of the zinc, it is necessary to use strictly pure zinc or else to amalgamate the zinc electrode with mercury to cover up the impurities.

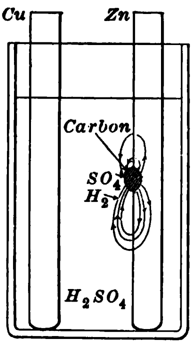


FIG. 225. — Local action in a cell.

The second defect is the fact that, when the poles of a simple cell are connected by a wire, the current does not remain constant, but rapidly gets weaker. This **polarization,** as it is called, is caused by the hydrogen bubbles which collect on the copper strip and thus form a gaseous coating. This layer of hydrogen is a poor conductor of electricity and therefore weakens the current. Furthermore the hydrogen layer has a slight battery action of its own, tending to send a current in a direction opposite to that desired, and this also weakens the current delivered by the cell.

Let us set up a zinc-sulphuric-acid-copper cell, connect it to a high resistance galvanometer, and observe the deflection. If we then short

circuit the poles of the cell by a short wire, which polarizes the cell quickly, we shall observe, on removing the wire, that the deflection is less than before. We may restore the cell by lifting the copper plate out of the acid for a moment or by brushing off the hydrogen bubbles.

We may also show polarization in a carbon-zinc cell in a similar way, but we can easily restore the cell by pouring into the acid a solution of potassium dichromate, a substance rich in oxygen. This increases the current because the hydrogen is taken up chemically by the oxydizing agent. If we now "short-circuit" the cell, that is, connect the terminals with a low-resistance conductor, the cell recovers quickly when the short circuit is removed. Such a substance as the potassium dichromate is called a **depolarizer**.

264. Commercial cells. There is a two-fluid cell, called the **Daniell cell**, which is free from polarization. In this cell the copper plate (Cu) stands in a solution of copper sulphate or blue vitriol ($CuSO_4$) and the zinc (Zn) in a solution of zinc sulphate ($ZnSO_4$). Both the copper sulphate and the zinc sulphate break up into ions. When the circuit is closed, both copper and zinc ions carry the current toward the copper electrode. The zinc ions, however, do not reach the copper plate, because zinc in copper sulphate replaces copper, forming zinc sulphate. The result is that the zinc goes into a solution forming zinc sulphate, and metallic copper is deposited on the copper electrode.

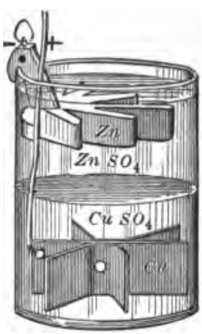


FIG. 226. — Gravity cell.

One form of this cell, much used in telegraphy, is called a **gravity cell** (Fig. 226) because the two liquids are separated by gravity. The dilute solution of zinc sulphate is lighter and therefore floats on the saturated solution of copper sulphate. The copper plate in the bottom of the jar is surrounded by crystals of copper sulphate to keep the solution saturated. In the dilute zinc-sulphate solution above is a heavy piece of zinc in the shape of a "crowfoot."

If the gravity cell is allowed to stand with its circuit open, the liquids mix slowly, and copper is deposited on the zinc in long festoons which cause local action, and sometimes grow long enough to short-circuit the cell. To prevent this, the external circuit must be kept closed. The cell is therefore well adapted for telegraphy, where a *small*, *constant* current is needed, but is not good for ringing doorbells or other *intermittent* work. In another form of Daniell cell, the solutions are separated by a cup of porous earthenware.

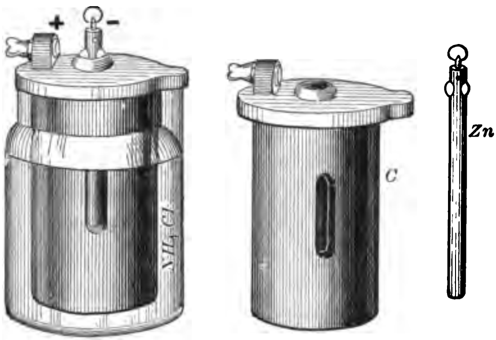


FIG. 227. — Sal-ammoniac cell.

For open-circuit work, such as ringing doorbells, the **sal-ammoniac cell** (Fig. 227) is used. The electrodes are zinc and carbon, and the electrolyte is a solution of sal-ammoniac (ammonium chloride, NH_4Cl). To reduce the polarization as much as possible, the carbon electrode is made with a large surface, and the cell often contains, as a depolarizer, a mixture of carbon and manganese dioxide. Since this depolarizer is slow in its action, the cell is adapted only to open-circuit work. It gives off no fumes, has very little local action, and so, when once set up, requires very little attention. Occasionally the water which has evaporated must be replaced and the zinc renewed.

The type of cell now most used for small intermittent work is the **dry cell**. This differs from the sal-ammoniac cell just described only in that the electrolyte is in the form of a paste instead of being a liquid. The negative electrode is the zinc can which contains the carbon and paste

(Fig. 228). The zinc is protected on the inside by several layers of blotting paper, and the space around the carbon is filled with a mixture of carbon, manganese dioxide, and sawdust, saturated with a solution of sal-ammoniac. The top is sealed with wax, and the whole cell is slipped into a pasteboard box.

FIG. 228. — Dry cell.

The dry cell is much used for ringing doorbells, running clocks, and operating the spark coils used to ignite gas engines on boats and automobiles. It requires no attendance, but *must not be left on closed circuit.* Sometimes the life of an exhausted dry cell can be extended slightly by punching a hole in the top and pouring in water, but usually exhausted cells are thrown away.

QUESTIONS

1. What are the points which a good cell should possess?
2. Why would you not use a gravity cell for ringing a doorbell?
3. What is the "fuel" in the dry cell?
4. Why are the small motors for fans, sewing machines, etc., never run by batteries if any other source of power is available?
5. If a person touches the poles of a cell, why does he not get a "shock"?
6. If you touch the two wires from a dry cell to the tip of your tongue, do you taste anything, and if so, why?

MAGNETIC EFFECT OF ELECTRIC CURRENT

265. Oersted's discovery. In 1819 a Danish physicist, Oersted, made a discovery which aroused the greatest interest because it was the first evidence of a connection between magnetism and electricity. He found that if a wire connecting the poles of a voltaic cell was held over a compass needle, the north pole of the needle was deflected toward the west

when the current flowed from south to north, as shown in figure 229, while a wire placed under the compass needle caused the north end of the needle to be deflected toward the east.

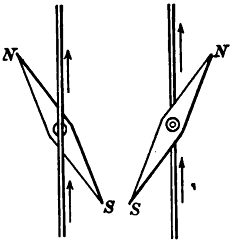


FIG. 229. — Deflection of magnetic needle by electric current.

266. Magnetic field around a current. Inasmuch as the compass needle indicates the direction of magnetic lines of force, it is evident from Oersted's experiment that a current must set up a magnetic field at right angles to the conductor. To make this clear, the student may perform the following experiment.

We will send a strong current down a vertical wire which passes through a horizontal piece of cardboard. To indicate the magnetic lines of force, we will sprinkle iron filings on the cardboard and tap it gently while the current is on. The filings arrange themselves in concentric rings about the wire. By placing a small compass at various positions on the board, we see that the direction of these lines of force is as shown in figure 230.

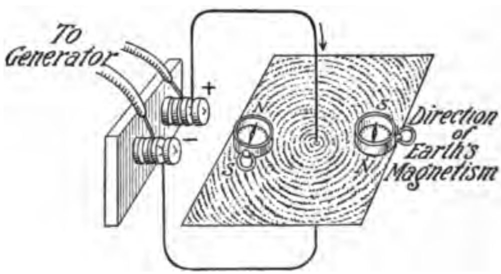


FIG. 230. — Magnetic lines of force around a current.

A convenient rule for remembering the direction of the magnetic flux around a straight wire carrying a current is the so-called **thumb rule**.

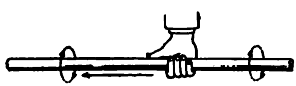


FIG. 231. — Thumb rule for magnetic field around a wire.

If one grasps the wire with the right hand (Fig. 231) *so that the thumb points in the direction of the current, the fingers will point in the direction of the magnetic field.*

If we know the direction of the magnetic field near a conductor, we can, by applying this rule, find the direction of the current.

Figure 232 shows the field around the wire with the current going *in* and figure 233, with the current coming *out*.

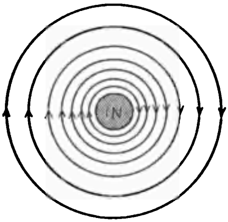


Fig. 232. — Current going *in*, clockwise field.

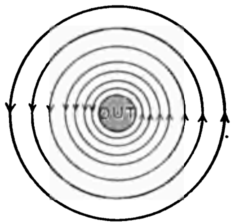


Fig. 233. — Current coming *out*, anticlockwise field.

267. Magnetic field around a coil. If a wire carrying a current is bent into a loop, all the lines of force enter the loop at one face and come out at the other face. If several loops are put together to form a coil, practically all the lines will thread the whole coil and return to the other end outside the coil.

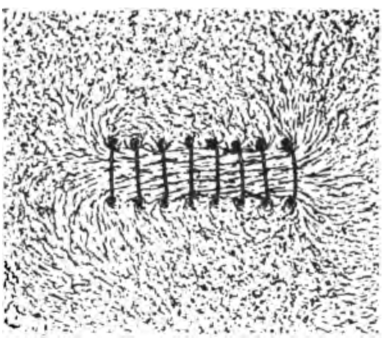


Fig. 234. — Magnetic flux around a coil.

(1) We may thread a loose coil of copper wire through a board or sheet of celluloid in such a way that when iron filings are evenly scattered over the smooth surface of the board, while a strong current is sent through the wire, they will indicate the lines of magnetic force (Fig. 234). By tapping the board gently and using a small compass, we can see the general direction of the lines of magnetic flux. It will be noticed that there are a few circular lines around each wire, and that these lines go out between the loops. They are called the "leakage flux" of the coil.

(2) If we send a current through a close-wound coil of insulated copper wire, and bring it near a compass needle, we find that it behaves like a bar magnet. If the current is reversed, the poles of the coil are reversed.

(3) If we put a soft iron core inside the coil when the current is on, the iron exerts a very strong pull on bits of iron; but when the current is off, the iron loses this magnetism almost at once.

(4) If we use a large horseshoe electromagnet, or a model of a magnetic hoist, and considerable current, we may show that a tremendous force can be exerted by an electromagnet.

An iron core in a coil of wire is so much more permeable than air that the same current in the same coil produces several thousand times as many lines of forces in the iron core as it would in air alone.

268. Electromagnet. An iron core, surrounded by a coil of wire, is called an electromagnet. It owes its great utility not so much to the fact of its great strength, as to the fact that, if it is made of soft iron, *its magnetism can be controlled at will.* Such an electromagnet is a magnet only when current flows through its coil. When the current is stopped, the iron core returns almost to its natural state. This loss of magnetism is, however, not absolutely complete; a very little **residual magnetism** remains for a longer or shorter time.

An electromagnet is a part of nearly every electrical machine, including the electric bell, telegraph, telephone, dynamo, and motor.

To determine its **polarity**, we shall find it convenient to express the thumb rule as used for a straight wire, in another way, as follows: —

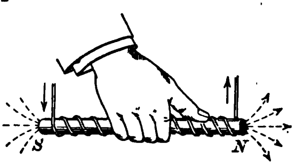


FIG. 235.—Rule for polarity of coil carrying current.

THUMB RULE FOR A COIL. *Grasp the coil with the right hand so that the fingers point in the direction of the current in the coil, and the thumb will point to the north pole of the coil* (Fig. 235).

The **strength** of an electromagnet depends on the strength of the current and on the number of loops or turns of wire.

MICHAEL FARADAY. Born in London, in 1791, the son of a blacksmith. Died in 1867. A chemist who made many wonderful discoveries in electricity and magnetism.

JOSEPH HENRY. Born in Albany, N. Y., in 1799. Died in 1878. Was for six years a schoolmaster at Albany Academy, for fourteen years a professor at Princeton, and for the rest of his life the head of the Smithsonian Institution in Washington. Made the first careful study of the electromagnet, and shares with Faraday the honor of discovering the laws of electromagnetic induction.

It is the practice, in order to make use of both poles of an electromagnet, to bend the iron core and the coil into the shape of a horseshoe, as shown in figure 236.

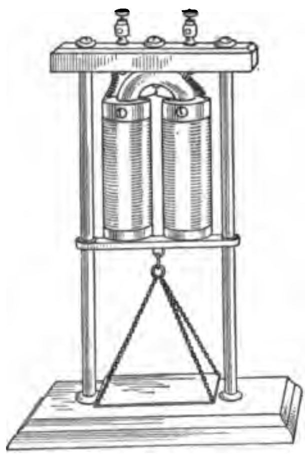


FIG. 236. — Electromagnet.

Practical electromagnets were made in 1831 by Joseph Henry, a famous American schoolmaster and scientist, then teaching in the academy at Albany, N.Y., and by Faraday in England. Henry's magnet was capable of supporting fifty times its own weight, which was considered very remarkable at the time.

Magnetic hoists are now built so powerful that when the face of the iron cores is brought in contact with iron or steel castings and the current is turned on, the magnets will lift from 100 to 200 pounds of iron per square inch of pole face, and yet release the load of iron the moment the current is cut off.

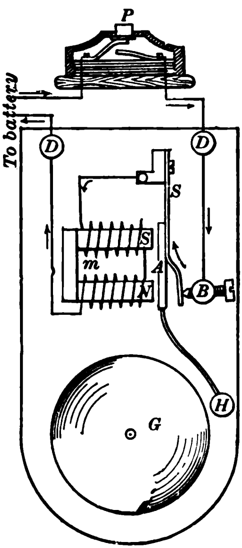


FIG. 237. — Electric-bell circuit.

APPLICATIONS OF THE ELECTROMAGNET

269. Electric bell. An electric-bell circuit usually includes a battery of two or more cells, a push button, and connecting wires, besides the bell itself (Fig. 237). When the circuit is closed by pushing the button *P*, the current flows through the electric magnet (*m*) and attracts the armature (*A*). As

the armature swings to the left, it pulls the spring (S) away from the screw contact (B) and breaks the circuit. This stops the current, and the electromagnet releases the armature. It then springs back again and closes the circuit at the screw, and the whole process is repeated. The swinging of the armature, which carries a hammer, causes a series of rapid strokes against the bell as long as the button is pushed. It does not matter in which direction the current flows. The construction of the push button P is shown more clearly in figure 238.

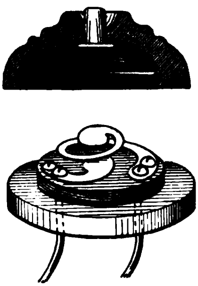


FIG. 238.—Push button.

270. What to do when the bell won't ring. First make sure that the connecting wires at the bell, push button, and battery are firmly screwed into the binding posts.

Next inspect the battery. The liquid should fill the jar within an inch of the top. The zinc should be clean and free from crystals and should dip into the solution, but should not touch the carbon.

If the battery consists of dry cells, you will do well to get a pocket ammeter and try each cell. A new cell will indicate about 20 amperes. If a cell has dropped much below 5 amperes, it is dead.

Next test the push button by removing the cover and holding a a piece of metal across the terminal wires. If the bell rings, it shows that the trouble is a poor connection in the button. Brighten up the contact points with sandpaper.

Finally look over the bell itself carefully, especially the point where the make and break occurs. Sometimes the screw with the platinum point gets loose or gets worn off and needs readjustment.

271. Telegraph. The word "telegraph" means an instrument which "writes at a distance," for the early forms invented by Samuel F. B. Morse, in 1844, were designed to make dots and dashes on a moving strip of paper. Nowadays the receiving instrument, called the sounder, makes a series of clicks separated by short or long intervals of time to represent the dots and dashes.

The **telegraph** consists essentially of a battery, a key, and a sounder, as shown in figure 239. Gravity cells are used in

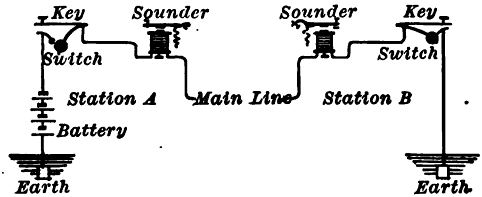


FIG. 239. — Simple telegraph circuit.

practical work, but for experimental purposes any kind of battery will serve.

The key (Fig. 240) is a device, something like a push button, for making and breaking the circuit. The sounder (Fig. 241) consists of an electromagnet with a soft iron armature which is fastened to a brass bar. This bar is pivoted so as to move up and down. When a current flows through the electromagnet, the armature is pulled down; when the circuit is broken, a spring pulls the bar up again. Two set screws above and below the bar limit its motion and make the clicks. As the clicks made by the bar hitting these two set screws are different, the ear recognizes the time between these two clicks as a dot or a dash according as the key is depressed a short or a long time.

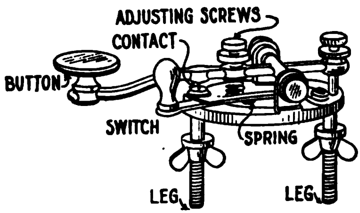


FIG. 240. — Telegraph key.

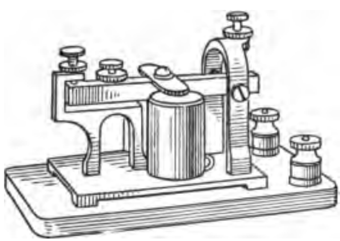


FIG. 241. — Telegraph sounder.

When the telegraph came into commercial use, it was found that the resistance of the connecting wires, called the

line, was so great that the current was too feeble to operate the sounder, even when many cells were connected in series. A relay (Fig. 242) is therefore employed to open and close the circuit of a local battery which operates the sounder. This relay contains an electromagnet whose coil has many turns of very small copper wire. In front of this magnet is a light iron lever which is held away from the electromagnet by a very delicate spring. The connections are shown in figure 243. When the key in the main circuit is closed, the weak current excites the relay magnet enough to pull the armature against a set screw, thus closing the local circuit which sends a strong current through the sounder.

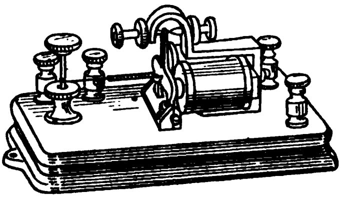


FIG. 242. — Telegraph relay.

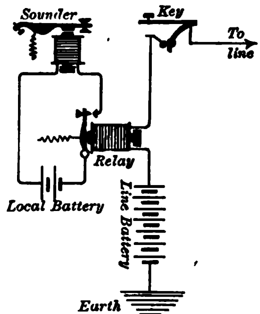


FIG. 243. — Diagram of relay telegraph circuit.

In ordinary telegraphy it is customary to use a single wire of galvanized iron or hard-drawn copper, and to use the earth as a return circuit. At each station along the line there is a local circuit consisting of battery and sounder, which is closed by a relay. The relay is in another circuit containing a key and the main-line battery or generator. Each key is provided with a switch so that the main circuit is kept closed everywhere except at the station where the operator is sending a message.

272. Other forms of telegraphs. Through the inventions of Edison and others we are now able to send two messages simultaneously in each direction. In other words, we can send four messages over a single wire all at the same time. This is called **quadruplex telegraphy.**

Submarine telegraphy began as early as 1837, but it was not till 1866 that a really successful Atlantic cable was laid. Such a cable contains a central conducting core of copper wires twisted together. This is surrounded by a thick insulating coating of rubber and outside of this is a protective covering of hemp and steel wires. The copper core and the steel sheath act like the coatings of an immense Leyden jar. The effect of this is to make the sending of messages very slow. The impulses received at the other end are also very weak. It was only when an exceedingly delicate receiving instrument had been devised by Lord Kelvin, that the first Atlantic cable could be used at all.

SUMMARY OF PRINCIPLES IN CHAPTER XV

Current flows *downhill*, from + to −, in *outside* circuit.
Current is pumped *uphill*, from − to +, *inside* of cell.

Energy is supplied by chemical action of acid on zinc.
Zinc is fuel of cell.

Current carried through cell by charged ions (pieces of molecules).

Lines of magnetic force around a straight current are concentric circles.
Thumb rule for *straight wire:* Use *right* hand. *Thumb* points with *current. Fingers* curl with *field.*

Lines of force around a coil mostly go through inside and come back outside.
Thumb rule for *coil:* Use *right* hand. *Thumb* points with *field* toward *N*-pole. *Fingers* curl with *current.*

QUESTIONS

1. An electromagnet is found to be too weak for the purpose intended. How may its strength be increased?

2. In looking at the *N* end of an electromagnet, in which direction does the current go around the core, clockwise or anticlockwise?

3. If you find that the *N*-pole of a compass held under a north and south trolley wire points toward the east, what is the direction of the current in the wire?

4. In a certain factory, steel was once used by mistake instead of soft iron to make the cores of the electromagnets for some bells. What would be the matter with the bells?

5. What would be the effect of winding an electromagnet with bare copper wire instead of insulated?

6. What sort of material is used to insulate copper wire which is to be used (*a*) to wind electromagnets, (*b*) to wire electric doorbell circuits, and (*c*) for electric lights?

7. What is the difference between a relay and a sounder that makes it possible for a weak current to work one and not the other?

CHAPTER XVI

MEASURING ELECTRICITY

Galvanometers — the ampere — ammeters — the ohm — internal and external resistance — the volt — voltmeters.

Ohm's law for whole circuit and for part of circuit — resistances in series and in parallel — cells in series and in parallel.

Specific resistance and the circular mil — effect of temperature on resistance — resistance boxes — measurement of resistance by voltmeter-ammeter method, and by Wheatstone bridge.

273. Necessity for a unit of current strength. In the construction, study, and use of electrical machinery, we are constantly dealing with electric currents. We say a current is strong or weak, just as we speak in a rough way of things being fast or slow, hot or cold. When, however, we go a step farther and ask *how strong* this current is or *how weak* that current is, we are forced to have some unit of current strength, and some means of measuring currents in terms of it.

274. Galvanometers. — We can get an idea of the *relative* strength of two currents by means of a galvanometer. There are two kinds of galvanometers in common use, the older of which is the **moving-magnet** galvanometer (Fig. 244). This consists of a compass needle pivoted or hung at the center of a large wooden frame on which are wound one or more turns of wire. This coil is set facing east and west so that the compass needle lies parallel to its plane. When a current is sent through the wire, an east and west magnetic field is set up at the center of the coil and the compass is deflected more or less according as the current is stronger or weaker.

In the type of galvanometer just described, the coil is large and is fastened firmly to the base, while the magnet is

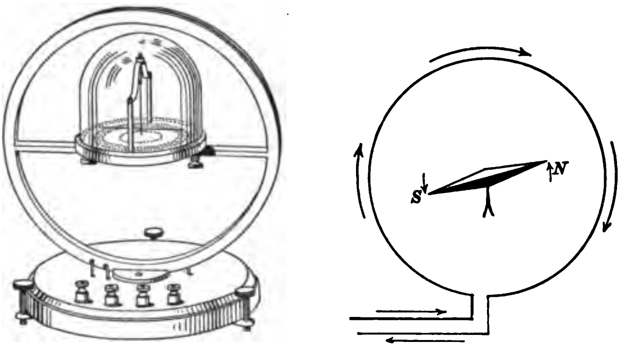


FIG. 244. — Moving-magnet galvanometer and diagram of essential parts.

small and movable. In the **moving-coil** type (Fig. 245), the magnet is large and is fastened firmly to the base, while the coil is small and movable. The magnet, *NS*, is usually made in the shape of a horseshoe so that it may be as strong as possible. The coil is wound on a very light rectangular frame and hangs between the jaws of the magnet. Usually there is a cylinder, *I*, of soft iron in the space inside the moving frame to still further increase the field. The bottom of the coil is connected with a binding post, *B*, by a spiral of very fine wire which carries the current into the coil without disturbing its freedom to rotate; the current leaves the coil through the fine suspension wire, *AC*. The top of this wire is twisted until the coil hangs in the plane of the poles *N*

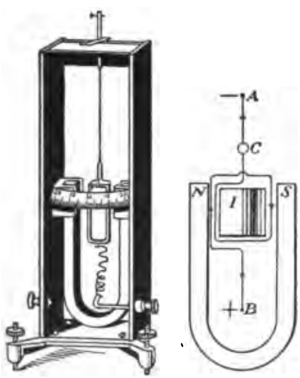


FIG. 245. — Moving-coil galvanometer and diagram of essential parts.

and S when no current is passing through it. If there is a current, the coil acts like a tiny magnet with poles pointing to the front and rear, and tries to turn itself so that these poles may get as near possible to the N and S poles of the magnet. The amount by which it is able to twist the suspension wire measures the current.

The moving-coil type is much more convenient for ordinary work, but the moving-magnet type can be made much more sensitive, and is used when very small currents have to be detected or measured.

275. The ampere. Having learned how to *compare* currents by means of a galvanometer, let us consider what the unit is, in terms of which currents can be *measured*. When we open the faucet at a sink, a current of water flows through the pipe. The rate of this flow can be easily measured in cubic feet (or gallons) of water per minute (or per second). Thus we speak of water as flowing at the rate of 1 gallon per second or 5 gallons per second. In much the same way we speak of electricity as flowing along a wire at the rate of 1 coulomb per second or 5 coulombs per second. The **coulomb** is the unit quantity of electricity, just as the gallon is the unit quantity of water. We have to consider the rate of flow of electricity so often that we have a special name for the unit rate of flow, **1 coulomb per second.** We call it an **ampere.** Thus 5 amperes means 5 coulombs per second.

It is possible to define the ampere in terms of the magnetic effect of an electric current, but, as a matter of fact, electrical engineers have agreed to define the ampere in terms of its chemical effect. If two silver (Ag) plates are placed in a jar of silver nitrate solution ($AgNO_3$), and if the + and − terminals of a battery are connected, one to one plate and one to the other, it will be found that the plate where the current goes *in* (the **anode**) loses in weight because silver is dissolved, and the plate where the current comes *out* (the **cathode**) gains in weight because silver is deposited. *By international agree-*

ment the quantity of electricity which deposits 0.001118 grams of silver is **one coulomb**, *and the current which deposits silver at the rate of 0.001118 grams per second is* **one ampere.** The apparatus used in the accurate measurement of current by this method is shown in figure 246. The anode is the silver disk at the left, and the cathode is the silver (or platinum) cup at the bottom. The porous cup at the right is put between the anode and the cathode in the solution like the cup in a Daniell cell.

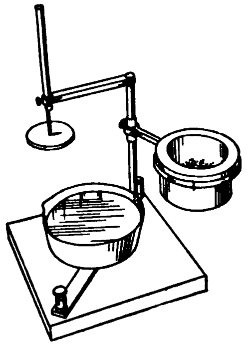


FIG. 246.—Silver coulombmeter.

276. Illustrations of the ampere. When a new dry cell is short-circuited with a short stout wire, about 20 amperes flow through the wire. An ordinary 16 candle power carbon filament electric lamp takes a little less than half an ampere, while the arc lamps used for street lighting require from 6 to 10 amperes. A telegraph sounder operates on 0.25 amperes, and a telephone receiver on less than 0.1 amperes, while the motor on a street car often takes as much as 40 or 50 amperes.

277. The ammeter. The legal method, described above, of defining an ampere is not, of course, a convenient method of measuring current strength. The coulombmeter is used only for standardizing the ammeters which are used in everyday life to indicate current strength.

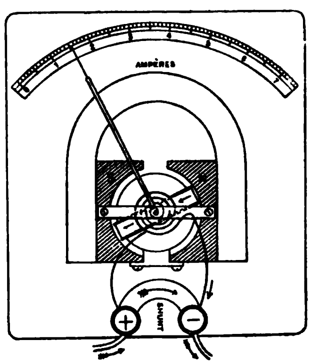


FIG. 247.—Ammeter.

The **commercial ammeter** (amperemeter) is a *shunted, moving coil galvanometer.* The instrument (Fig. 247) contains a coil of fine insulated copper wire, wound on a light frame, and mounted in jeweled bearings between the

poles of a strong permanent horseshoe magnet. A fixed soft iron cylinder midway between the poles of the magnet concentrates the field. The moving coil rotates in the gap between the core and the pole pieces. The coil is held in equilibrium by two spiral springs, which serve also to carry the current into and out from the coil. Only a small fraction, perhaps 0.001 of the current to be measured, goes through the movable coil, the major part being carried past the coil by a metal strip called a **shunt.** Since the current through the coil is a constant fraction of the whole current, the pointer which is attached to the moving coil can be made to indicate directly on a graduated scale the number of amperes in the total current.

It will be seen that the resistance of an ammeter, which is practically the resistance of the shunt, is very small, and that the whole current passes through the instrument.

278. Electrical resistance. Although we divide substances into two classes, conductors and non-conductors or insulators, yet even the best conductors of electricity are not perfect. This means that all conductors offer some **resistance** to the flow of electricity and transform a part of the energy which they carry into heat. We are already familiar with the fact that a stream of water flowing through a pipe is held back or retarded by the friction of the pipe. The amount of this friction depends on the smoothness of the inner surface, the length and the size of the pipe. So with electricity, the resistance of a conductor depends: —

(1) On the material used; iron, for example, has nearly 7 times as much resistance as copper;

(2) On the length; a wire 10 feet long has twice as much resistance as a wire 5 feet long;

(3) On the size of the wire; a wire 0.04 inches in diameter has one fourth the resistance of a wire 0.02 inches in diameter;

(4) On the temperature; heating a copper wire from 0° to 100° C increases its resistance about 40 %.

279. Legal ohm, the unit of resistance. The legal unit of resistance, called the **ohm**, is the resistance at 0° C of a column of mercury 106.3 centimeters long, with a cross section of about 1 square millimeter (more exactly, of uniform cross section and weighing 14.4521 grams). This legal definition of the ohm fixes the material, length, cross section, and temperature of the conductor whose resistance is taken as the standard. Since a column of mercury is not convenient to handle, we ordinarily use "standard coils" made of some high-resistance alloy, such as German silver or manganin.

280. Illustrations of the ohm. About 157 feet of # 18 copper wire (the size ordinarily used to connect electric bells), or 26 feet of iron wire or 6 feet of manganin wire of the same size, has a resistance of 1 ohm. The resistance of a small electric bell is about 3 ohms, of a telegraph sounder 4 ohms, of a relay 200 ohms, of a telephone receiver 60 ohms, and of 16 candle power incandescent lamp 220 ohms when hot.

281. Internal and external resistance. It must not be forgotten that there is resistance to the flow of electricity in every part of an electric circuit. In the case of the electric-bell circuit, there is the bell itself, the push button, the connecting wires, and the battery. The resistance of the generator, whether it be a battery or a dynamo, is called the **internal resistance**, and that of the rest of the circuit is the **external resistance**. It is the gradual increase in the internal resistance of a long-used dry cell which cuts down the current it can deliver and so destroys its usefulness.

282. Electromotive force. In hydraulics we know that to get water to flow along a pipe it is essential to have some driving or motive force, such as that furnished by a pump. In much the same way, to get electricity to flow along a wire we must have an **electromotive force**, such as that furnished by a battery or dynamo. The unit of electromotive force is the **volt**. *A volt may be defined as the electromotive force needed to drive a current of one ampere through a resistance of one ohm.*

283. Illustrations of volts. A common dry cell gives about 1.4 volts, and a storage cell about 2 volts. A gravity cell gives about 1.08 volts, and the so-called Weston Standard cell, used in very exact voltage measurements, gives 1.0183 volts at 20° C. The current for lighting a building is usually delivered at 110 or 220 volts, and street cars operate on about 550 volts.

284. Distinction between volts and amperes. The intensity of an electric current is measured in **amperes**, the electromotive force driving the current is measured in **volts**. In a given circuit the greater the electromotive force is, the greater is the current. We know that we must have a certain "head" of water in order to get a given number of gallons of water to flow through a given pipe each second; so we must have a certain electromotive force to make a given current of electricity flow through a given wire. With both water and electricity we must have a motive force in order to have a current, but we may have the motive force and yet have no current. If the valve is closed in the water pipe or the switch is open in the electric circuit, we might have motive force (volts) but no current (amperes).

COMPARISON OF HYDRAULIC AND ELECTRICAL UNITS

UNITS	WATER	ELECTRICITY
Quantity	Gallon	Coulomb
Current	Gallon per sec.	Ampere = 1 coulomb per sec.
Motive force	"Feet of head"	Volt
Resistance		Ohm

285. The voltmeter. The commercial **voltmeter** is simply a *high-resistance galvanometer*. When electromotive force is applied to a galvanometer, the current it allows to pass is proportional to the voltage, and so the scale can be graduated to read the voltage directly. The instrument (Fig. 248) is usually a moving coil galvanometer, like an ammeter. Indeed the same instrument is often used for both purposes. A voltmeter does not have a shunt between its terminals,

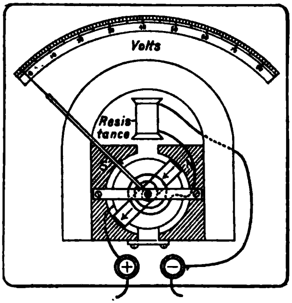


FIG. 248. — Voltmeter.

like an ammeter, but does have a large resistance coil inserted in series, so that only a very small current passes through the instrument, but all of it goes through the moving coil. In fact such a voltmeter gives correct values only when the current used is so small as not to affect appreciably the voltage to be measured.

This will be understood by considering the water analogy shown in figure 249. It is evident that the current in the connecting pipe *AB* is a good measure of the difference in level between *L* and *L′*, only when the current in *AB* is so small as not to change appreciably the levels whose difference is to be measured.

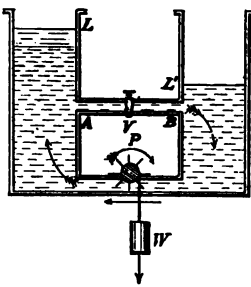


FIG. 249. — Water analogue of a voltmeter.

To make voltmeters usable over different ranges we have merely to connect coils of different resistance in series with the same galvanometer.

Since the voltmeter is an instrument for measuring the electromotive force between the *two ends* of a circuit or of part of a circuit, it must have its terminals connected to the two points; that is, it must be put *across* the circuit, not *in* it. The proper connections for both ammeter and voltmeter are shown in figure 250.

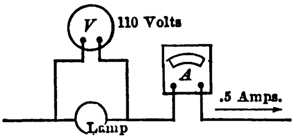


FIG. 250. — How to connect a voltmeter and ammeter.

286. Electromotive force of a cell. If the electromotive force of a simple cell is observed with a voltmeter, it will be found that the voltage of the cell is not changed by moving the plates near together or far apart, or by lifting them almost out of the liquid so as to change greatly their effective size. These changes affect the current sent by the cell through an external circuit by changing the *internal resistance* of the cell, not its voltage.

If a stick of carbon is used instead of a copper plate, the voltage of the cell will be found to be greater; if, however, hydrochloric acid is used instead of sulphuric, the voltage is less.

All this shows that the voltage of a cell depends, not on its size, but on the materials of which it is made. A very large storage cell with plates several square feet in area, such as is used in power stations, gives exactly the same 2 volts that a tiny cell in a test tube would if made of the same materials. The large cell can drive more current through a given circuit than a small one because its internal resistance is so small. Of course, with constant use, the small cell would be exhausted much more quickly.

287. Ohm's law. Let a Daniell cell be connected in series with a considerable resistance, perhaps 100 ohms, and a galvanometer, and the current noted. If two cells are used in series, the current will be about twice as great. If, without changing the number of cells, we double the external resistance, the current will be about half as great. If we halve the external resistance, the current will be doubled.

In general, we find that the current increases as the electromotive force increases, and that the current decreases as the resistance in the circuit increases. A German physicist, Ohm (1789–1854), was the first to state this relation between current, electromotive force, and resistance. The law is: *The intensity of the electric current along a conductor equals the electromotive force divided by the resistance.*

$$\text{Current} = \frac{\text{electromotive force}}{\text{resistance}}.$$

In electrical units: —

$$\text{Amperes} = \frac{\text{volts}}{\text{ohms}}.$$

In symbols: —

$$I = \frac{E}{R},$$

where I = Intensity of current in amperes,
E = Electromotive force in volts,
R = Resistance in ohms.

If we know the current and resistance and want the electromotive force, we have

$$E = IR.$$

If we know the electromotive force and current and want to calculate the resistance, we have

$$R = \frac{E}{I}.$$

288. Examples using Ohm's law.

1. What is the intensity of the current sent through a resistance of 5 ohms by an electromotive force of 110 volts?

$$I = \frac{E}{R} = \frac{110}{5} = 22 \text{ amperes.}$$

2. What electromotive force is needed to send a current of 0.03 amperes through a resistance of 1000 ohms?

$$E = IR = 0.03 \times 1000 = 30 \text{ volts.}$$

3. Through what resistance will 110 volts send a current of 10 amperes?

$$R = \frac{E}{I} = \frac{110}{10} = 11 \text{ ohms.}$$

PROBLEMS

1. Find the intensity of the current which an electromotive force of 10 volts sends through a resistance (*a*) of 3 ohms, (*b*) of 40 ohms.
2. How much electromotive force is needed to send 2 amperes through (*a*) 2 ohms, (*b*) 50 ohms?
3. What is the resistance of a circuit when the electromotive force is 110 volts and the current intensity is 2 amperes?
4. An electric heater of 10 ohms resistance can safely carry 12 amperes. How high can the voltage run?
5. An electromagnet draws 4 amperes from a 110-volt line. How much would it draw from a 220-volt line?
6. A certain dry cell has an electromotive force of 1.5 volts and will give about 27 amperes when short-circuited. What is its internal resistance? What is the internal resistance of the same cell when, after much use, it will give only 9 amperes?

289. Application of Ohm's law to partial circuits. Not only does Ohm's law apply to an entire circuit, the current in the entire circuit being equal to the total electromotive force divided by the resistance of the *entire* circuit, but it also applies to any *part* of a circuit. That is, the current in a certain part of a circuit equals the voltage across that same part divided by the resistance of the part.

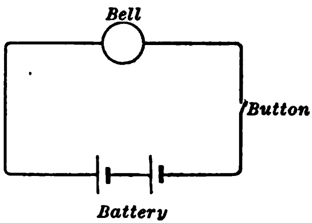


FIG. 251. — Bell circuit.

For example, suppose the electromotive force of a battery is 3 volts, the resistance of the bell (Fig. 251) is 3 ohms, the resistance of the wires and button is 1.5 ohms, and the internal resistance of the battery is 0.5 ohms. To find the intensity of the current, we have

$$I = \frac{E}{R} = \frac{3}{3 + 1.5 + 0.5} = 0.6 \text{ amperes.}$$

The current has the same intensity throughout the circuit. To find the voltage across the bell, we have

$$E = IR = 0.6 \times 3 = 1.8 \text{ volts.}$$

To find the voltage drop within the battery, we have

$$E = IR = 0.6 \times 0.5 = 0.3 \text{ volts.}$$

Since the total electromotive force of the battery is 3 volts, and it takes 0.3 volts to send the current through the battery itself, the *terminal voltage* of the battery is 3 – 0.3, or 2.7 volts. Of this 1.8 volts is needed to send the current through the bell and the remainder, 0.9 volts, is used to send the current through the connecting wires and push button.

If a voltmeter were connected across the battery, it would read 2.7 volts, or the *terminal voltage*. The total electromotive force (e. m. f.) is computed by multiplying the current in the circuit, 0.6 amperes, by the total resistance, 3 + 1.5 + 0.5 = 5 ohms.

$$E = IR = 0.6 \times 5 = 3.0 \text{ volts total e. m. f.}$$
$$3.0 - 0.3 = 2.7 \text{ volts terminal voltage.}$$

290. Terminal voltage of a cell depends on its current. Connect a voltmeter to the terminals of a dry cell, and note the e. m. f. Then connect a coil of very high resistance (1000 ohms) across the ter-

minals. The terminal voltage, as indicated by the voltmeter, is very nearly the same as before. But if we connect a short, thick wire across the terminals, so as to draw a large current, we see by the voltmeter that the terminal voltage drops instantly.

Thus we see that the voltage drop in a cell depends directly upon the current used, and that the terminal voltage decreases when the current increases.

291. Series arrangement. Let us consider still further the electric current in apparatus arranged **in series**.

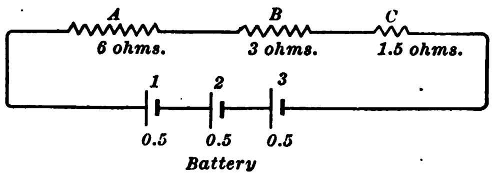


Fig. 252. — Three cells and three resistances in series.

In figure 252 we have three cells and three resistances connected in series. By this we mean that the carbon of cell 3 is joined to the zinc of cell 2, and the carbon of cell 2 is joined to the zinc of cell 1. The circuit then runs from the carbon of cell 1 through the resistances A, B, and C in succession, back to the zinc of cell 3.

The laws governing series circuits are as follows: —

*The **current** in every part of a series circuit is the same.*

*The **resistance** of several resistances in **series** is the **sum** of the separate resistances.*

*The **voltage** across several resistances in **series** is equal to the **sum** of the voltages across the separate resistances.*

Moreover, since the voltage is equal to the resistance times the current ($E = IR$), and since the current (I) in every part of a series circuit is the same, it follows that the *voltage across any part of a series circuit is proportional to the resistance of that part.*

For example, in figure 252, if the e. m. f. of each cell is 2 volts, the e. m. f. of the three cells in series is 3 times 2, or 6 volts.

If the resistance of A is 6 ohms, of B, 3 ohms, of C, 1.5 ohms, and of each cell, 0.5 ohms, the total resistance is $6 + 3 + 1.5 + (3 \times 0.5) = 12$ ohms.

The current is $\frac{6}{12}$, or 0.5 amperes.

The voltage across A is 6 times 0.5, or 3 volts, across B, 3 times 0.5, or 1.5 volts, and across C, 1.5 times 0.5, or 0.75 volts.

The voltage "drop" in each cell is 0.5 times 0.5, or 0.25 volts, so that the terminal voltage of each cell is $2 - 0.25$, or 1.75 volts.

292. Parallel arrangement. When the several resistances are so arranged that the current divides between them, as shown in figure 253, part going through A, part through B, and the rest through C, they are said to be **in parallel** or **multiple** or **shunt.**

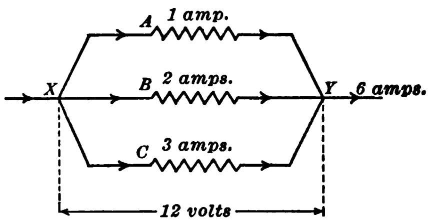


FIG. 253. — Three resistances in parallel.

The *voltage* across each separate resistance of the three parallel resistances is the *same*. For example, if the voltage across A is 12 volts, then the voltage across B is 12 volts, and also across C, for each resistance lies between the same two points, X and Y.

The currents, however, in each of the resistances in parallel are *not the same*, unless the resistances are all equal. If, for example, the resistances of A, B, and C are each 6 ohms, the current in each is 2 amperes. But if the resistances are unequal, the *greatest* current will flow through the *smallest* resistance. Of course the *total current* passing through a parallel arrangement of resistances is equal to *the sum* of the currents in the separate conductors. Thus if the current in A is 1 ampere, in B, 2 amperes, and in C, 3 amperes, the total current through the combination, and through the rest of circuit, is $1 + 2 + 3$, or 6 amperes.

293. Calculation of resistances in parallel. If we know the voltage and *total* current through a set of resistances in parallel we have merely to apply Ohm's law to compute the

total resistance of the combination. Thus, if the voltage across A, B, and C in figure 253 is 12 volts, and the total current is 6 amperes, the total resistance is 2 ohms.

In case we know only the separate resistances and want to compute their combined resistance in parallel, we may well make use of the idea of **conductance.** By conductance we mean the **ease** with which a current flows along a wire, while **resistance** represents the **difficulty.**

Conductance is the reciprocal of resistance; that is,

$$\text{Conductance} = \frac{1}{\text{resistance}}.$$

The unit of conductance is the **mho**, which is the unit of resistance (ohm) spelled backwards.

$$1 \text{ mho} = \frac{1}{1 \text{ ohm}}.$$

Thus a wire has a conductance of 1 mho when its resistance is 1 ohm, and a conductance of 2 mhos when the resistance is 0.5 ohms.

It can be easily shown that *the conductance of a parallel arrangement is equal to the sum of the conductances of the separate parts.*

Thus suppose three resistances, A, B, and C, are arranged in parallel and are 12, 6, and 4 ohms respectively. Find their combined resistance.

Let R = resistance of A, B, and C in parallel.

Then $\frac{1}{R} = \frac{1}{12} + \frac{1}{6} + \frac{1}{4}.$

Whence $R = 2$ ohms.

294. Cells arranged in parallel. Not only may the separate *external* resistances be arranged in parallel, but the *cells* or generators themselves may be so arranged. For example, in figure 254 we have a battery of three cells arranged in parallel.

The laws governing such a case are as follows: —

The voltage of the battery is the voltage of one cell.

The internal resistance of the battery is the resistance of one cell divided by the number of cells, since three cells, for example, have three times the conductance of one cell.

The current will be the sum of the currents through each cell.

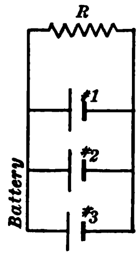


FIG. 254. — Three cells in parallel.

For example, suppose that in figure 254 the voltage of each cell is 2 volts, the resistance of each cell is 0.6 ohms, and the external resistance R is 0.3 ohms, and we desire to find the current through each cell.

The total current I is found by Ohm's law; thus

$$I = \frac{E}{R} = \frac{2}{0.3 + \frac{0.6}{3}} = \frac{2}{0.5} = 4 \text{ amperes.}$$

Then the current in each cell will be $\frac{1}{3}$ of 4 amperes or 1.3 amperes.

295. Best arrangement of cells of a battery. By means of a lecture table voltmeter we may show that the e. m. f. of 6 cells in series is 6 times that of one cell, and that the e. m. f. of 6 cells in parallel is the same as that of one cell.

By using an ammeter and a small external resistance, we can show that 6 cells arranged in parallel give more current than 6 cells in series; but with considerable external resistance, the series arrangement furnishes the greater current.

In general, to make the intensity of the current as large as possible, when the **external resistance is large**, arrange the **cells in series**, but when the **external resistance is small**, arrange the **cells in parallel**.

We can easily see the reason for this if we recall from the foregoing discussion that in a series battery the voltage and

the internal resistance of each cell are both multiplied by the number of cells. In the parallel arrangement of cells the voltage is not increased, and the internal resistance of each cell is divided by the number of cells. From Ohm's law we have

$$I = \frac{E}{R + r},$$

where R is the external resistance and r the internal. If R is much larger than r, it does not make much difference just how large r is, and we should make E as large as possible by arranging the cells in series. But if R is small, r becomes the important part of the denominator, and it pays to make r as small as possible by arranging the cells in parallel.

In practical work the external resistance is usually large compared with the internal resistance, so the cells of a battery are generally arranged in series.

PROBLEMS

1. Three resistances, $A = 80$ ohms, $B = 60$ ohms, and $C = 40$ ohms, are put in parallel and the voltage across the combination is 120 volts. Find the current in each, the total current, and the resistance of the combination.

2. If a battery is used to light 20 lamps arranged in parallel, and each lamp requires 0.5 amperes, how many amperes must the battery supply?

3. What e. m. f. will be needed to force 2 amperes through a series circuit, containing a battery of resistance $\frac{1}{2}$ ohm, a line of resistance 1 ohm, and a lamp of resistance 100 ohms?

4. Three lamps of 150 ohms each are joined in series. If each lamp requires 0.5 amperes, what is the total current required? What is the total voltage required?

5. If the three lamps of problem 4 were arranged in parallel, what would be the total current and total voltage needed?

6. Six cells, each having an e. m. f. of 2 volts and an internal resistance of 0.3 ohms, form a series battery to send a current through a resistance of 50 ohms. How strong is the current?

7. If the cells of problem 6 are arranged in parallel, what is the current strength?

8. Calculate the current strength sent by the six cells of problem 6, arranged in series, through an external resistance of 0.1 ohms. Also the current when the cells are in parallel.

9. If a current of 0.25 amperes is needed in a telegraph circuit, how many gravity cells in series will be required, if each has an e.m.f. of 1.08 volts and a resistance of 2 ohms, and if the line resistance is 500 ohms?

10. Six cells are arranged 3 in series and 2 in multiple (Fig. 255) to send a current through an external resistance of 4 ohms. If each cell has an e.m.f. of 1.5 volts and a resistance of 0.5 ohms, how intense will the current be?

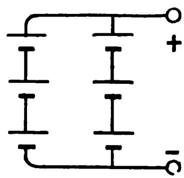


FIG. 255. — Six cells, three in series, and two in multiple.

11. A galvanometer, whose resistance is 299 ohms, has a short stout wire of 1 ohm resistance connected across the terminals. What fraction of the total current goes through the galvanometer?

12. A storage battery is sending current through two wires in parallel, each having a resistance of 10 ohms. If the current through the battery is 6 amperes, what is the voltage drop in each wire?

13. A wire of 4 ohms is connected in series with 2 wires joined in parallel and having resistances of 8 and 12 ohms. Find the total resistance.

14. A dry cell when tested with a voltmeter showed 1.5 volts, and when tested with an ammeter whose resistance was negligible, gave 7.5 amperes. Find the internal resistance of the cell.

15. If the voltage drop in a trolley line carrying 150 amperes is 12.5 volts, what is the resistance of the line?

296. Computation of resistance. For the measurement of voltage and of intensity of current we have direct reading voltmeters and ammeters, but as yet we have no simple instrument for measuring resistance directly in ohms. In many cases, however, we can compute the resistance of a wire, if we know its material, length, size, and temperature.

Since wires are usually round, it is inconvenient to compute their area of cross section in square inches. Consequently electrical engineers call a wire, which is one thousandth of an inch in diameter, **1 mil** in diameter and its area of cross section **1 circular mil.** Inasmuch as the areas of circles vary as the squares of their diameters, *the*

area of a wire expressed in circular mils is equal to the square of its diameter expressed in mils.

For example, suppose a wire is 0.015 inches in diameter. What is its cross section in circular mils?

0.015 inches = 15 mils; area = $(15)^2 = 225$ circular mils.

The resistance of a wire is usually computed by comparing it with the resistance of a wire of the same material and of a standard size and length. The standard usually chosen is 1 foot long and 1 circular mil in area of cross section. Such a piece of wire is called a **mil foot** of wire.

The resistance of a mil foot of wire is sometimes called the **specific resistance** of the substance of which the wire is made. For example, the specific resistance of copper is about 10.4 ohms, of aluminum 17.4 ohms, and of iron 64 ohms.

Since the resistance of a conductor varies directly as its length and inversely as its area of cross section, we can readily compute the resistance of a wire when its length and diameter are given.

Suppose we wish to find the resistance of 500 feet of #18 copper wire. Since the specific resistance of copper is 10.4, we know that the resistance of 1 mil foot of copper wire is 10.4 ohms. So that of 500 feet of copper wire 1 mil in diameter is 500×10.4, or 5200 ohms. But from the wire tables given on page 304, we find that #18 wire is 40.3 mils in diameter. Therefore its cross section is $(40.3)^2$, or 1624 circular mils. Therefore the resistance will be $\frac{1}{1624}$ of the resistance of a wire 1 mil in diameter, or

$$\tfrac{1}{1624} \times 5200 = 3.2 \text{ ohms}.$$

As this computation has to be made very often in practical work, it is convenient to put it in the form of an equation.

$$R = \frac{kl}{d^2},$$

where R = resistance in ohms,
k = specific resistance (ohms per mil foot),
l = length in feet,
d = diameter in mils.

Thus $$R = \frac{10.4 \times 500}{(40.3)^2} = 3.2 \text{ ohms}.$$

297. Effect of temperature on resistance. If we coil about 10 feet of #30 iron wire around a piece of asbestos and send a current through it, we can observe with a lecture-table ammeter that, as we heat the wire in a Bunsen flame, the intensity of the current is greatly reduced.

If we connect an incandescent lamp in series with a coil of iron wire, as shown in figure 256, we can observe by the dimming of the lamp that the current becomes less when the iron wire is heated.

Experiments show that the resistance of 1 mil foot of copper wire at 20° C or 68° F is 10.4 ohms, while at 0° C it is 9.6 ohms. The resistance of a one-ohm coil of copper, correct at 0° C, increases as the temperature rises, approximately 0.0042 ohms for each degree. For example, a coil which measures 10 ohms at 0° C will at 50° C have a resistance of $10 + (0.0042 \times 50 \times 10) = 12.1$ ohms. By carefully measuring the resistance of a wire when cold and then when hot, we have an electrical method of measuring temperature.

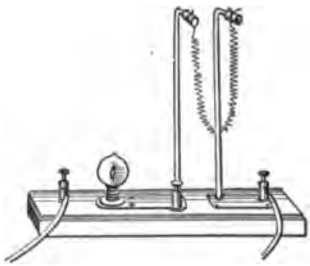


FIG. 256.—Iron wire when hot has more resistance than when cold.

Most pure metals have nearly the same rate of increase of resistance with rise of temperature. Most alloys of metals not only have a much higher resistance than the pure metals of which they are made, but are much less affected by temperature changes. For example, "manganin" is an alloy of copper, nickel, and iron manganese, which has a specific resistance of from 250 to 450 ohms according to the proportion of the metals used, but its resistance shows scarcely any change with temperature.

There are a few substances, such as carbon, glass, and porcelain, which *decrease* in resistance when heated. For example, the resistance of a carbon filament lamp when it is hot is about half of the resistance of the same filament when it is cold.

PROBLEMS

Find the resistance of each of the wires described in problems 1 to 6: —

1. One mile, # 10 copper wire, diameter 0.102 inches.
2. Fifty feet, # 16 copper wire, diameter 0.051 inches.
3. Twenty feet, # 30 copper wire, diameter 0.010 inches.
4. Two miles, # 14 iron wire, diameter 0.064 inches.
5. Two hundred feet, # 10 iron wire, diameter 0.102 inches.
6. Five thousand feet, # 6 aluminum wire, diameter 0.162 inches.
7. Find the number of feet of # 20 iron wire needed to make a resistance of 5 ohms.
8. Find the diameter of a copper conductor which has a resistance of 2 ohms per 1000 feet.
9. What size of copper wire must be used for a trolley wire 4 miles long, if the line resistance must not exceed 2 ohms?
10. What is the "line drop," that is, voltage drop, in a 4-mile copper wire carrying 100 amperes, if the wire is 0.325 inches in diameter? (Voltage drop $E = IR$.)

298. Rheostats and resistance boxes. To control an electric current, we must regulate either the voltage or the resistance. As electricity is usually supplied to us at fixed voltages, such as 110, 220, or 500 volts, we have to control the intensity of the current by a variable resistance, called a **rheostat.** For example, in starting a motor, for reasons which will be discussed in Chapter XVIII, the current must not be thrown on at full intensity at first, and so a rheostat (Fig. 257) is inserted. By moving a lever arm the resistance is gradually cut out as the motor comes up to speed. Rheostats are usually made of some high-resistance alloy such as German silver, or of carbon (lamps), or sometimes of water with a little salt dissolved in it.

It is not enough to know the resistance of a rheostat. We must know also its carrying capacity, for the electrical energy consumed in the rheostat is converted into heat and must be radiated off as fast as it is produced. Otherwise the temperature will rise to a dangerous point, so that the wire melts or sets on fire things which are near it.

A **resistance box** is also made of resistance coils, but since they are used for electrical measurements which involve only

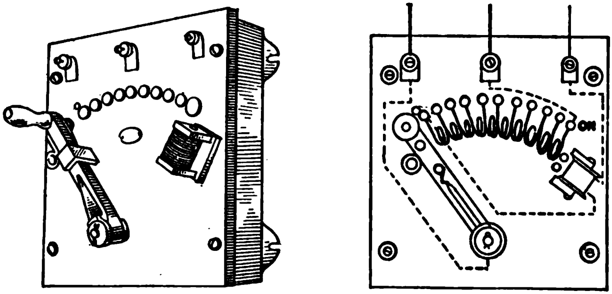


FIG. 257. — Starting rheostat.

small currents, they have very small carrying capacity. The coils have definite resistances, such as 1, 2, 3, 5, 10, 20 ohms, and are made of a wire which is only slightly affected by temperature changes. The resistance box corresponds for electrical measurements to a set of weights used in weighing.

For convenience, the coils are usually mounted in a box, as shown in figure 258, which has an insulating hard rubber

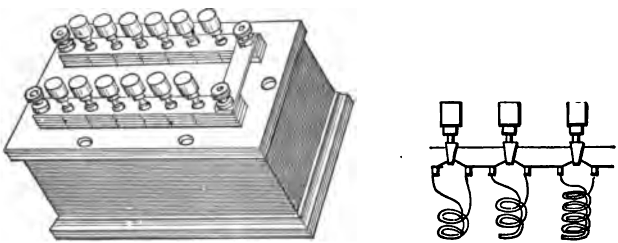


FIG. 258. — Resistance box.

top. On this are fastened a series of brass blocks which can be connected by brass plugs which fit between them. Inside the box are the various coils wound on spools. The ends of a coil are connected to adjoining blocks, so that each gap is

bridged inside by a coil. At each end of the series of blocks is a terminal binding post. When all the plugs are *firmly* in place, the only resistance is that of the series of blocks and of the plugs themselves, which is usually negligible; but when a certain plug is removed, the resistance of that coil is introduced.

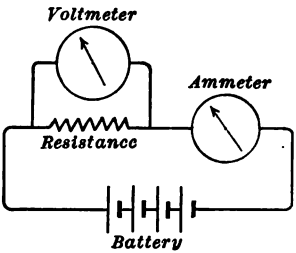


FIG. 259. — Resistance measured by a voltmeter and an ammeter.

299. Measurement of resistance by voltmeter-ammeter method. As has been said, there is no simple instrument for measuring a resistance directly, as the voltmeter measures voltage, or an ammeter current. But there are two ways of measuring a resistance indirectly.

If extreme accuracy is not required, the method shown in figure 259 is commonly used. The unknown resistance is placed in series with an ammeter, and the voltage across the resistance is obtained by a voltmeter. Then, by Ohm's law,

$$R = \frac{E}{I}.$$

It is essential that the resistance of the voltmeter be so high that practically no current goes through it. This method also requires that both ammeter and voltmeter be accurately calibrated; that is, compared with standard instruments and the errors noted.

300. Measurement of resistance by Wheatstone bridge. A more accurate method of measuring resistance is the **Wheatstone bridge,** which is a machine for balancing resistances. It consists essentially of a loop of four resistances, R, X, m, and n, arranged as in figure 260. When the key (K) is closed, the current from the cell flows into the loop at A, and there divides so that part (I_1) goes through AC and part (I_2) through AD. A sensitive galvanometer is con-

nected between C and D. Then the resistances R, m, and n are so adjusted that no current flows through the galvanometer, which means that all of I_1 has to go on through CB and all of I_2 through DB, and also that C and D are "equipotential" points. When this adjustment has been made, the voltage drop across AC is I_1R, and the voltage drop across AD is I_2m. But since C and D are at the same potential, these voltage drops are equal, and

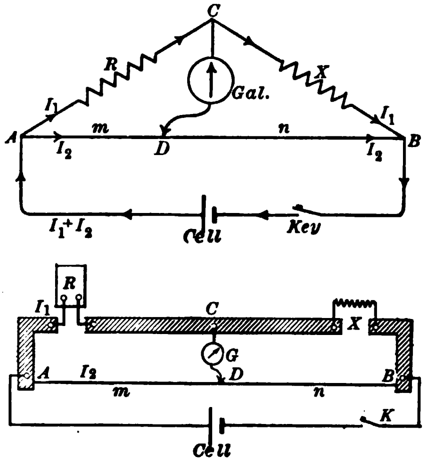


FIG. 260. — Wheatstone bridge to balance resistances.

$$I_1R = I_2m. \qquad (1)$$

For similar reasons

$$I_1X = I_2n. \qquad (2)$$

Dividing equation (1) by equation (2), we have

$$\frac{R}{X} = \frac{m}{n}.$$

From this fundamental equation of the Wheatstone bridge, if we know R, m, and n, we can compute X.

In one form of this apparatus the resistance ADB consists of a wire of uniform cross section one meter long. Since the resistances m and n are then directly proportional to the distances AD and DB, the equation becomes

$$\frac{R}{X} = \frac{\text{Distance } AD}{\text{Distance } DB},$$

Wire Tables

American or Brown and Sharp (B. and S.) Gauge

Gauge	Diameter in Mils	Area in Circular Mils	Diameter in Millimeters	Carrying Capacity, Rubber Insulation, Amperes
0000	460.	211,600.	11.68	210.
000	410.	167,800.	10.40	177.
00	365.	133,100.	9.27	150.
0	325.	105,500.	8.25	127.
1	289.	83,690.	7.35	107.
2	258.	66,370.	6.54	90.
3	229.	52,630.	5.83	76.
4	204.	41,740.	5.19	65.
5	181.9	33,100.	4.62	54.
6	162.0	26,250.	4.12	46.
7	144.3	20,820.	3.67	—
8	128.5	16,510.	3.26	33.
9	114.4	13,090.	2.91	—
10	101.9	10,380.	2.59	24.
11	90.7	8,234.	2.31	—
12	80.8	6,530.	2.05	17.
13	72.0	5,178.	1.83	—
14	64.1	4,107.	1.63	12.
15	57.1	3,257.	1.45	—
16	50.8	2,583.	1.29	6.
17	45.3	2,048.	1.15	—
18	40.3	1,624.	1.02	3.
19	35.4	1,288.	.90	
20	32.0	1,022.	.81	
21	28.5	810.	.72	
22	25.3	643.	.64	
23	22.6	509.	.57	
24	20.1	404.	.51	
25	17.90	320.	.46	
26	15.94	254.	.41	
27	14.20	202.	.36	
28	12.64	159.8	.32	
29	11.26	126.7	.29	
30	10.02	100.5	.26	
31	8.93	79.7	.23	
32	7.95	63.2	.20	
33	7.08	50.1	.18	
34	6.30	39.7	.16	
35	5.61	31.5	.14	
36	5.00	25.0	.13	
37	4.45	19.83	.11	
38	3.96	15.72	.10	
39	3.53	12.47	.09	
40	3.14	9.89	.08	

where R is a known resistance, such as a resistance box, and the distances AD and DB are read off on a meter stick. It may help one to remember this equation to observe that

$$\frac{\textbf{Left resistance}}{\textbf{Right resistance}} = \frac{\textbf{Left distance}}{\textbf{Right distance}}.$$

For example, suppose R is 5 ohms and AD is 45.5 centimeters; then DB is 54.5 centimeters, and

$$\frac{5}{X} = \frac{45.5}{54.5}.$$

$$X = 6.05 \text{ ohms.}$$

Problems

1. Compute the resistance of a lamp through which a voltage of 113 volts sends a current of 0.4 amperes.
2. Find the resistance of a street-car heater which takes 5 amperes of current from a 550 volt line.
3. A wire 50 feet long has a drop of 2 volts across it. Find the drop across 20 feet.
4. In a slide-wire Wheatstone bridge, the known resistance is 12 ohms, and the balance is obtained when AD (Fig. 260) is 42.5 centimeters. Compute the value of the unknown resistance.
5. In testing a Wheatstone bridge, 4 ohm and 6 ohm coils are inserted in the loops AC and CB. Find the position which D should have on the meter wire ADB.

SUMMARY OF PRINCIPLES IN CHAPTER XVI

Unit of current is ampere. Corresponds to gallons per second.
Unit of resistance is ohm. Corresponds to friction in pipe.
Unit of e. m. f. is volt. Corresponds to "head."

Ammeter — low resistance — put in series — carries whole current.
Voltmeter — high resistance — put across circuit — diverts small current.

E. M. F. of cell = total pump action in cell.
Terminal voltage = potential difference between terminals.

Terminal voltage less than e. m. f. by amount needed to keep current moving through internal resistance of cell.

Ohm's law: —

$$\text{Current} = \frac{\text{e. m. f.}}{\text{resistance}}.$$

Applies to whole circuit, or to any part of circuit.

If applied to whole circuit, must take account of *internal resistance* of cell, as well as of *external resistance.*

For resistances in series: —

Current everywhere the *same.*

Resistance of combination is *sum* of resistances of parts.

Voltage across combination is *sum* of voltages across parts.

For resistances in parallel: —

Total current through combination is *sum* of currents through parts.

Conductance of combination is *sum* of conductances of parts.

Voltage across conductors *same* for all.

For cells in series: —

E. m. f. is sum of e. m. f.'s of parts.

Resistance is sum of resistances of parts.

Current same in all cells as in external circuit.

For cells in parallel: —

E. m. f. is same as e. m. f. of one cell.

Resistance of n cells in parallel is $\frac{1}{n}$ th the resistance of any one alone.

Current in each cell is $\frac{1}{n}$ th the current in external circuit.

$$\text{Resistance of wire} = \frac{\text{specific resistance (mil foot)} \times \text{length (feet)}}{\text{square of diameter (mils)}}.$$

In slide-wire Wheatstone bridge: —

$$\frac{\text{Left resistance}}{\text{Right resistance}} = \frac{\text{left distance}}{\text{right distance}}.$$

QUESTIONS

1. Why are telegraph lines usually made of iron wire, while trolley wires are made of copper?

2. Why should the circuit of a dry cell be kept open when the cell is not in use?

3. Why should a gravity cell be left on closed circuit when not in use?

4. What would happen if an ammeter were connected across the line? (Don't try it!)

5. What would happen if a voltmeter were put in series in a line?

6. Why is the moving-magnet type of galvanometer inconvenient?

7. What is the use of the shunt in an ammeter?

8. A copper wire and an iron wire of the same length are found to have the same resistance. Which is the larger?

9. Why do we get a more intense current by moving the plates of a cell close together?

10. What is the effect on the current strength of allowing the liquid to evaporate to half its volume in a sal-ammoniac cell, and why?

11. What instruments would we need in addition to a coulombmeter to measure the intensity of a current?

12. Why should keys be inserted in both the battery line and the galvanometer line of a Wheatstone bridge?

13. Why are electric bells usually arranged in parallel instead of in series?

CHAPTER XVII

INDUCED CURRENTS

Induction by permanent magnets — direction of induced current — induction by electromagnets — induction coil — jump spark ignition — self-induction — make and break ignition — telephone.

301. Faraday's discovery. If we had to depend on batteries for all our electric currents, we should not be lighting our streets and houses with electric lamps or riding on electric cars. The cost of zinc as a fuel in the voltaic cell makes the battery too expensive as a source of large quantities of electricity.

It is, however, possible to turn mechanical energy directly into electrical energy by means of a machine called a dynamo, in which currents are induced by moving magnets. It was the discovery of the dynamo that made possible the modern age of electricity.

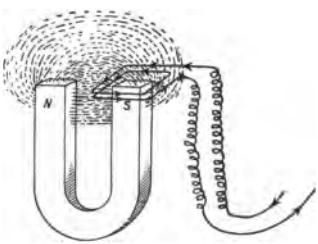


FIG. 261. — A coil moving in a magnetic field generates a current.

302. Currents induced by magnets. If we connect the ends of a coil of many turns of fine insulated wire to a lecture-table galvanometer, and then move the coil quickly down over one pole of a strong horseshoe magnet, as shown in figure 261, we observe a deflection. When we raise the coil again, we observe a deflection in the opposite direction. If we lower the coil again and hold it down, we find that the galvanometer pointer comes back to zero. If we repeat the experiment, moving the coil down slowly and up slowly, we find that the deflection is less than before.

Such experiments show that it is possible to produce momentary electric currents without a battery. An electric current produced by moving a coil in a magnetic field is called an **induced current.** It is evident from the experiment that the current is induced only when the wire is **moving** and that the direction of the current is reversed when the motion changes direction. Since an electric current is always made to flow by an electromotive force, the motion of a coil in a magnetic field must generate an **induced electromotive force.** Experiments have shown that this induced electromotive force varies directly as the **speed** of the moving coil.

303. Direction of induced currents. If we take the same apparatus (Fig. 261) and move the coil down over the *N*-pole of the magnet and then down over the *S*-pole, we find that the deflections are in opposite directions in the two cases. To determine in which direction the induced current is flowing in the coil, one may make a little voltaic cell by putting in his mouth a copper wire and a zinc wire connected to the galvanometer. Since we know that the copper is the positive electrode, we can compare the direction of the galvanometer deflection caused by the cell current with that caused by the induced current, and so determine the direction of the latter. In this way we find that when the coil is moving down over the *N*-pole of the magnet, the induced current is in such a direction that the lower face of the coil is an *N*-pole. In a similar way we find that when the coil is brought down over the *S*-pole of the magnet, the induced current is in such a direction that the lower face of the coil is an *S*-pole. In both cases the lower face of the coil is a pole of such a sort as to be repelled by the pole toward which it is moving.

The direction of induced currents may be stated as follows: *An induced current has such a direction that its magnetic action tends to resist the motion by which it is produced.*

304. Currents induced by currents. Since an electromagnet can be made more powerful than a steel magnet, we would expect greater induced currents when we move an *electromagnet* near a coil.

We will connect the secondary coil *S* in figure 262 to a galvanometer and the primary coil *P* to a battery. When we move the current-carry-

ing primary coil *P* either into or out of the other coil *S*, a current is induced, just as when we move a magnet in and out of a coil. The induced current is, however, much greater. We find also that a stronger current in the coil *P* increases the strength of the magnetic field, and so of the induced current.

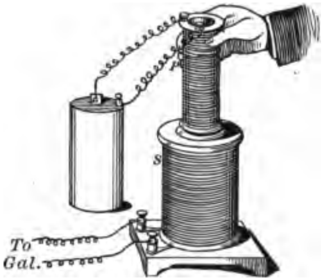


FIG. 262.—A moving electromagnet generates a current.

We may also increase the induced currents by inserting an iron core inside the primary coil. This greatly strengthens the magnetic field and so increases the number of lines of force about the coil.

If we put the primary coil with its iron core inside the secondary coil, we can generate an induced current by opening and closing a switch in the primary circuit. When the switch is opened and closed, the deflections are in opposite directions.

In general we see that an *induced current is set up in a coil whenever there is a change in the number of lines of magnetic force passing through the coil.*

QUESTIONS

1. Show how a current can be produced in a coil of wire by the motion of a magnet.
2. Why is it necessary in the experiment just described to use a coil of many turns?
3. Show how a coil of wire should be rotated in the earth's magnetic field to get the maximum induced current.
4. Show how a coil of wire can be rotated in the earth's magnetic field so as to get no induced current.

305. Induction coil. In the induction coil (Fig. 263) the core *c* is made of soft iron wires; the primary coil *P* consists of a few turns of large copper wire, and the secondary coil *s*, which is carefully insulated from the primary, contains many turns of very small silk-covered copper wire. To make and break the primary current very rapidly, an interrupter *H* is commonly made to operate on the end of the coil. This

automatic make and break works exactly like the electric bell described in section 269. When the primary circuit in such a coil is broken, the current tends to keep on as if it had inertia, and may jump the switch gap at *A* even after it has opened slightly. This slows up the "break" and weakens

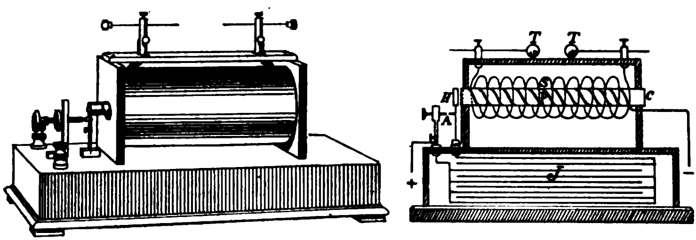


FIG. 263. — Induction coil.

the induced e.m.f. So a condenser *J* is connected across the gap. It is usually made of sheets of tin foil, insulated by paraffin paper, arranged as shown in figure 263. This furnishes a storage place into which the current can surge when broken, and diminishes the sparking at the interrupter. Even with a condenser there is some sparking, and so the contact points have to be tipped with silver or platinum and frequently cleaned.

Coils are generally rated according to the distance between the terminals of the secondary across which a spark will jump. When the coil is in operation, sparks jump across this gap in rapid succession, provided the terminals are close enough together. This type of coil is sometimes called the **Ruhmkorff coil** and is the one in general use for **jump-spark ignition** on gas engines.

306. Uses of induction coils. Jump-spark ignition is the most important practical application of the induction coil. Small induction coils are also used under the name of **medical** or **household coils.** These are usually so made that the strength of the induced current in the secondary can be

varied either by moving the primary coil in and out of the secondary, or by moving in and out a brass tube which fits around the core. In recent years very large and powerful induction coils have been built for **exciting X-ray tubes** and for setting up electric waves for **wireless telegraphy**. These uses will be described in Chapter XXIV.

307. Self-induction. It is a familiar fact in mechanics that bodies act as if disinclined to change their state, whether of rest or motion, and we call this tendency **inertia**. We have found a similar electrical phenomenon, when the primary of an induction coil is broken. Let us examine it more closely.

If an electric circuit contains a coil of wire with many turns and with a soft iron core, it opposes the building up of a current at the start, and when the circuit is broken, the current, once started, tries to keep on flowing, as shown by the spark at the gap. This electromagnetic inertia of a circuit is called its **self-induction**.

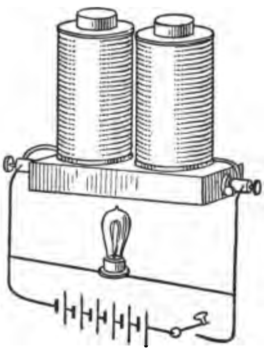


FIG. 264. — Self-induction of an electromagnet.

To show self-induction, we may put a small lamp across the terminals of a large electromagnet. If we throw on some supply of direct current, as from a storage battery, as shown in figure 264, the lamp lights up at first, but quickly dies down when the current becomes steady. When the switch is opened, the lamp again lights up.

In this experiment, when the circuit is closed, the self-induction of the coil opposes the flow of current through the electromagnet, and so the current has to go through the lamp. When the circuit is opened, self-induction causes the current to continue to flow, and the lamp is its only available path. *Self-induction, then, occurs only when the current is changing.*

308. Applications of self-induction. The principle of self-induction is made use of in **make-and-break ignition**. A *single*

coil is used, consisting of many turns of wire wound on a soft iron core. When such a coil is employed to furnish a spark in the cylinder of a gas engine, the circuit is as shown in figure 265. The terminals are two points inside the cylinder of the engine, one stationary (*A*) and the other moving (*B*). When *A* and *B* separate, the self-induction of the coil causes enough induced e.m.f. to make a spark jump across the gap between them.

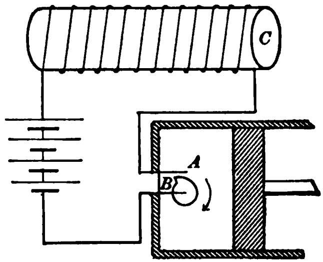


FIG. 265. — Make-and-break spark coil.

This is the kind of coil which is used to light gas burners in houses by means of a battery current. If the circuit of a large electromagnet, such as the field of a dynamo, is broken, while one is touching the conductors on either side of the gap, the current due to self-induction sometimes gives a severe shock.

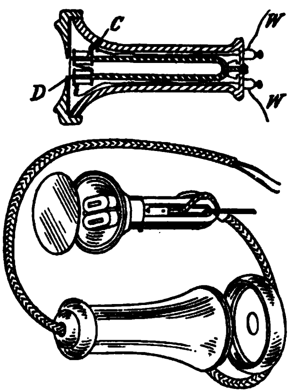


FIG. 266. — Bell telephone receiver.

309. Telephone receiver. In 1876 Alexander Graham Bell astonished the world by showing that the sound of the human voice could be transmitted by electricity. The essential part of his apparatus was what we still use and know as the **Bell receiver.** The hard rubber case contains a steel U-shaped magnet, which has around each pole a coil of many turns of very fine wire (Fig. 266). A disk of thin sheet iron is so supported that its center does not quite touch the ends of the magnet. A hard rubber cap or earpiece with a hole in the center holds the disk in place.

To show the operation of the telephone receiver, we will connect a receiver, in series with a lamp, to the A.C. mains or to a magneto, which furnishes an alternating current. We immediately hear a loud hum. If we hold the receiver upright and stand a pencil on the diaphragm, it dances up and down. The alternating current, sent through the coil, alternately opposes and strengthens the magnet, which attracts the disk alternately more and then less, thus causing it to vibrate. This sets the air to vibrating and produces sound.

310. The microphone. To understand how the right sort of currents can be produced to make a telephone receiver speak words instead of merely humming, we will set up an old fashioned instrument called a microphone.

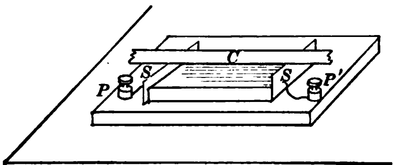


FIG. 267.—Carbon microphone.

A simple microphone can be made out of three lead pencils, or three pieces of electric light carbon, or out of a single carbon resting across two old safety razor blades (Fig. 267). If such a microphone is connected in series with a battery and telephone receiver, and a watch is laid on its baseboard, the ticks can be heard in the telephone even if it is some distance away. The little jars which the watch gives the baseboard shake the carbons so that the resistance at their points of contact varies and thus changes the current. The changing current then pulls the telephone diaphragm back and forth, and sets the surrounding air in motion.

311. The telephone transmitter. The modern "**solid back**" **telephone transmitter** is simply a carefully designed microphone. It contains a little box *C* (Fig. 268) which is filled with granules of hard carbon. The front and the back of the box are polished plates of carbon, and the sides of the box are insulators. The front carbon is attached to the center of the diaphragm *D*, and moves in and out a little when the diaphragm vibrates. The other plate is fastened

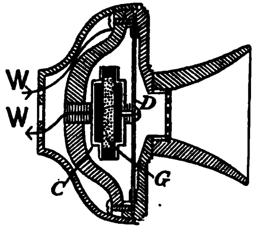


FIG. 268.—Carbon transmitter.

rigidly to the solid back of the case. A current from a battery flows through the diaphragm to the front plate, then back through the granules to the other plate, and then out along the telephone line to a receiver. When the diaphragm moves back a little, it compresses the granules, their resistance decreases, and the current gets stronger and pulls the diaphragm of the receiver back also. When the transmitter diaphragm moves out, the current decreases and the receiver diaphragm moves out also. So all the motions of the transmitter diaphragm are reproduced by the receiver diaphragm. If one speaks into the transmitter, causing its diaphragm to move in a corresponding way, the receiver diaphragm moves in the same way and produces the same kind of waves in the surrounding air.

312. Central *vs.* local batteries. The system we have just described is the one in use in all large cities. The battery is a large storage battery (or a dynamo) at the central station and is used on all the lines that happen to be busy at any instant.

In many country exchanges and on isolated lines another system, called the local battery system, is used because it is cheaper to install and maintain. Even in cities something equivalent to this system is used in "long-distance" work.

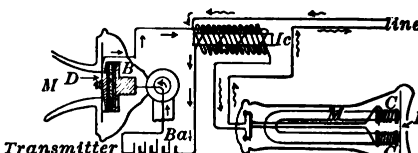


Fig. 269. — Local battery telephone system.

In this system (Fig. 269) each subscriber's telephone set contains a few dry cells which are connected in series with his transmitter, as already described. But the varying current produced, instead of being sent directly out on to the line, goes to the primary of a little induction coil and back to the battery. The secondary of the induction coil meanwhile sends out into the line an induced current that varies

exactly like the primary current, but is at much higher voltage. This, as we shall see in Chapter XVIII, makes the "line losses" much smaller, and so more energy gets through to the receiver than if the original current had been transmitted directly.

This system is really better, electrically, than the central battery system (Fig. 270). It is not used in large cities, chiefly because of the trouble involved in keeping so many local batteries in proper working condition.

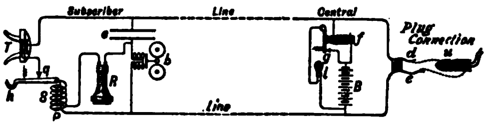


FIG. 270. — Telephone with central battery.

313. Return wire necessary. Telephone circuits used to be made like telegraph circuits, with only one wire, the return being through the earth. But in cities this is impractical because of the noise and confusion caused by stray currents in the earth due to trolley cars and other electrical disturbances. So in cities two wires are used, and they are put close together so that no currents can be induced in them by stray magnetic fields from other circuits (which would cause "cross-talk"), or from lighting and power circuits.

SUMMARY OF PRINCIPLES IN CHAPTER XVII

Induced current exists *only* when the number of lines of force through the circuit is *changing*.

Induced current has such a direction as to *oppose* the *motion* that causes it.

Self-induction appears *only* when the current is *changing*.

The effect of self-induction is always to *oppose* the *change* of the current.

QUESTIONS

1. Why is it that the self-induction of a circuit is not apparent as long as the current is steady?

2. Why is the lamp described in the experiment in section 307 not bright all the time that the switch is closed?

3. Why is it dangerous to touch the terminals of the secondary of a large Ruhmkorff coil?

4. What is likely to happen to an induction coil if you short-circuit the secondary while the coil is running?

5. Why is the induced e. m. f. in the secondary of an induction coil so much greater at the break of the primary than at the make?

6. What furnishes the energy of an induced current?

7. Upon what three factors does the e. m. f. of an induced current depend?

8. In a central battery telephone system, what arrangements are made to keep the battery from sending current all the time through telephone lines not in use?

CHAPTER XVIII

ELECTRIC POWER

The generator — wire cutting lines of force — dynamo rule for direction of current — law of induced e. m. f. — revolving loop commutator — Gramme ring — drum armature — field excitation.

Electric power — how measured — the joule and watt — commercial units.

Electric heating — common applications — fuses — computation.

Electric lighting — the arc — modern forms — incandescent lamps — metal filaments — efficiency.

The motor — side push on wire carrying current — motor rule for direction of motion — forms of motor — back e. m. f. — starting box — applications — efficiency.

Chemical effects — electrolysis — bleaching — electroplating — electrotyping — refining metals — electrochemical equivalents — storage battery.

THE GENERATOR

314. The importance of the generator. The most useful application of induced currents did not come until nearly forty years after Henry and Faraday made their wonderful discovery. Then the generator was developed, by means of which the enormous energy of steam engines and water wheels can be transformed into electricity. The electricity generated in this way can be transmitted many miles, and used in motors to turn all sorts of machinery, in lamps of various kinds to light our streets and homes, in heaters to warm cars and sometimes houses, to toast bread and heat flatirons, and in furnaces to melt steel in iron mills. Thus

the generator has revolutionized modern industry by furnishing cheap electricity.

315. Wire cutting lines of magnetic force. A simple way to get at the fundamental idea of the generator is to think, as Faraday did, of the induced e. m. f. produced in a *single wire when it is moved through a magnetic field.* Suppose the straight wire AB is pushed down across the magnetic field shown in figure 271. An induced e. m. f. is set up in AB, which makes B of higher potential than A, as can be shown by connecting B and A with a galvanometer. As long as the wire remains stationary no current flows. Even if the wire does move, if it moves in a direction parallel to the lines of force, no current flows. In short, a wire, to have an e. m. f. induced in it, must *move so as to cut lines of force.*

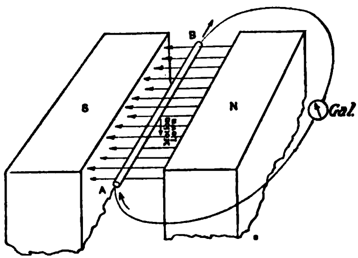


FIG. 271. — Induced e. m. f. in a wire cutting lines of force.

316. Direction of induced e. m. f. We have just seen that when the wire AB in figure 271 is moved down, the induced current in it is from A to B. If the wire were moved up, the induced current would be from B to A. Furthermore, if the field is reversed without changing the direction of motion of the wire, the current reverses. It will be seen, then, that the direction of the induced e. m. f. depends upon two factors, (a) the direction of the motion of the wire and (b) the direction of the flux or magnetic lines of force. The relation of these three directions may be kept in mind by Fleming's rule of three fingers, as shown in figure 272.

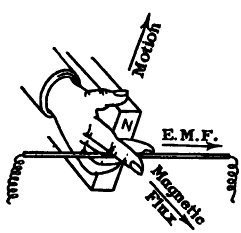


FIG. 272.—Right-hand rule for induced e. m. f.

FLEMING'S RULE. *Extend the thumb, forefinger, and center finger of the right hand so as to form right angles with each other. If the **thumb** points in the direction of the **motion** of the wire, and the **forefinger** in the direction of the magnetic **flux**, the **center** finger will point in the direction of the induced **current**.*

To remember this rule, notice the corresponding initial letters in the words "fore" and "flux," "center" and "current."

317. Amount of induced e. m. f. If we have a large electromagnet with flat-faced pole pieces (Fig. 273), we can demonstrate the various laws about induced currents in a conductor. If we move a wire down through the gap between the pole pieces, a millivoltmeter will show the induced current. If we hold the wire at rest in the gap, we observe no current. If we move the wire horizontally parallel to the lines of magnetic flux, we get no current. If we move the wire up through the gap, we observe a current in the opposite direction, as predicted by Fleming's rule. If we increase the magnetic field by increasing the current through the electromagnet, we increase the induced current. If we move the wire more quickly through the gap, we increase the induced current. Finally, if we bend the wire into a loop of several turns, and move the loop down over one pole so that all the wires on one side of the loop pass through the gap, we find that the current is increased.

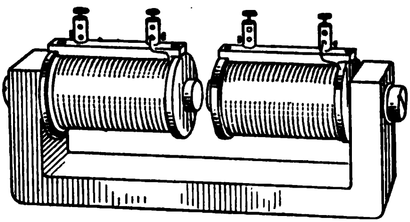


FIG. 273.—Electromagnet for demonstrating induced e. m. f.

In this experiment we see that the induced e. m. f. is increased by moving the wire faster across the magnetic field, by making the magnetic field stronger, and by using more turns of wire. In short, *the amount of induced e. m. f. depends on three factors:* (1) *the speed;* (2) *the magnetic field;* and (3) *the number of turns.*

Experiments show that

Induced e. m. f. varies as speed × flux × turns.

318. Commercial generators. A machine for converting mechanical energy into electrical energy is called a **dynamo** or **generator.** Its essential parts are two, (1) the **magnetic field,** which is produced by permanent magnets, as in the magneto, or by electromagnets, as in larger generators, and (2) a **moving coil** of copper wire, called the **armature,** wound on a revolving iron ring or drum. The armature wires correspond to the moving wires in the experiments above.

319. Current in a revolving loop of wire. If we rotate a rectangular coil between the poles of a large horseshoe magnet, or better, of an electromagnet, we can detect an electric current in the revolving coil by connecting it with flexible leads to a galvanometer. As we turn the coil, the current is reversed every half revolution.

It will help us to understand just what is happening in the revolving coil if we first consider what would happen in a single loop of wire which is rotated in a magnetic field, as shown in figure 274. If we start with the plane of the loop vertical and turn the handle in a clockwise direction, the wire *BC* moves *down* during the first half turn, and so, by Fleming's rule, we should expect the induced e. m. f. to tend to send the current from *C* to *B*. At the same time the wire *AD* is moving *up*, and the current will tend to flow from *A* to *D*. The result is that during the first half turn the current goes around the loop in the direction *ADCB*. During the second half turn the current is reversed and goes around in the direction *ABCD*.

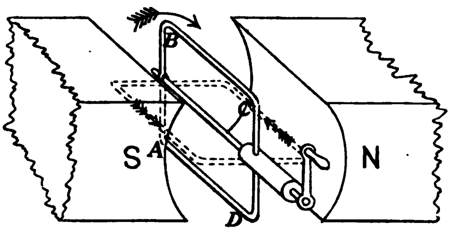


FIG. 274. — Single loop of wire turning in a magnetic field.

To show that this really does happen in the loop, we can cut the wire and connect the ends to *slip rings* *x* and *y*, as

shown in figure 275. The brushes B' and B'', which rest on the rings, are connected to a galvanometer. In this way it can be shown that there is generated in the coil an **alternating current** which reverses its direction twice in every revolution. Moreover, it is possible to show that the induced e.m.f. starting at zero goes up to a maximum and then back to zero in the first half turn; then it reverses and goes to a maximum in the opposite direction and finally back to zero. The induced e. m. f. reaches its maximum when the coil is horizontal, because in this position the wires AD and BC are cutting lines of force most rapidly. This is illustrated by the curve shown in figure 276.

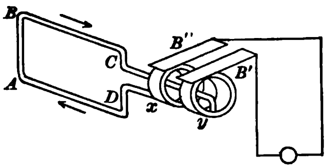


Fig. 275.—Single loop connected to slip rings.

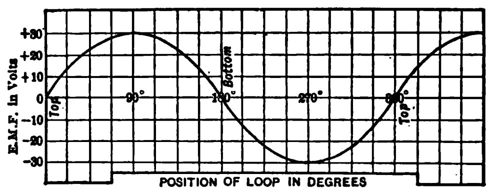


Fig. 276. — Curve to show relation of induced e.m. f. to position of loop.

Machines which are built to deliver alternating currents are called **alternators** or **A. C. generators.**

320. Commutator. To get a **direct current,** that is, one which flows always in the same direction, we have to use a **commutator.** To understand how this works, let us study a very simple case. If the ends of the loop in section 319 are connected to a **split ring,** as shown in figure 277, we may set the brushes $B+$ and $B-$ on opposite

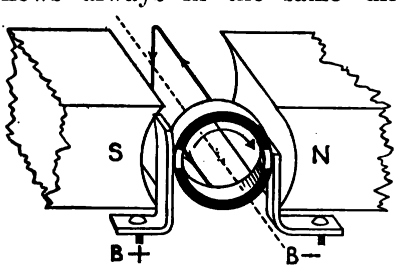


Fig. 277. — Split-ring commutator.

sides of the ring, so that each brush will connect first with one end of the loop and then with the other. By properly adjusting the brushes, so that they shift sections on the commutator just when the current reverses in the loop, that is, when the loop is in a vertical position, we may get the current to flow only *out* at one brush $B+$, and only *in* at the other brush $B-$. The direction of the current in the external circuit is always the same, even though the current in the loop itself reverses twice in every revolution.

The current delivered by such a machine can be represented by the curve in figure 278. Although it is always in the same direction, it is pulsating.

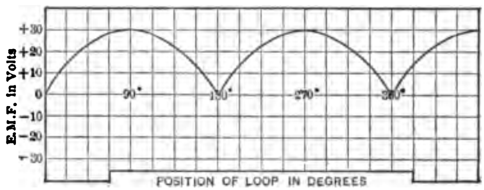


FIG. 278. — Pulsating e. m. f. delivered by loop fitted with commutator.

A machine with a commutator for delivering direct current is called a **direct-current dynamo** or **D.C. generator.**

321. Generators of steady currents. The e.m.f. produced by rotating a single loop in a magnetic field can be raised by using many turns of wire and by rotating the coil very fast. Nevertheless the current will be pulsating, and this is unsatisfactory for many purposes. To get a machine to deliver a *steady* current, a Frenchman, named Gramme, invented in 1870 the so-called **Gramme ring** form of armature.

The Gramme ring armature is now very seldom used, but it is worth studying carefully because the fundamental principles of its action can be understood from very simple diagrams, whereas most armatures of the common or drum type, although based on exactly the same principles, cannot be represented by simple diagrams.

A rotating soft-iron ring or hollow cylinder is mounted be-

tween the poles of an electromagnet, as in figure 279. The ring serves to carry the flux across from one pole to the other. There are scarcely any lines of force in the space inside the ring. A continuous coil of insulated copper wire is wound on the ring, threading through the hole every turn. When the ring rotates, as in figure 280, the wires on the *outside* are cutting lines of force, but those inside are not. Furthermore, according to the right-hand rule, the outside wires on the right-hand side are moving in such a direction that the induced current tends to flow towards us. The wires lying on the other side of the ring are moving so as to induce a current away from us. If there were no outside connections, these two opposing e. m. f.'s would just balance, and no current would flow. This would be like arranging a lot of cells in series with an equal number turned so that they are opposed to the first group [Fig. 281 (*a*)]; obviously no current would flow.

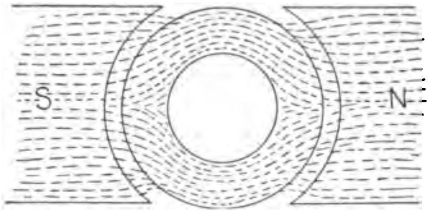


FIG. 279. — Magnetic field in a Gramme ring.

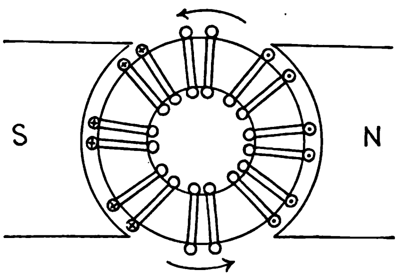


FIG. 280. — Gramme ring rotating in a magnetic field.

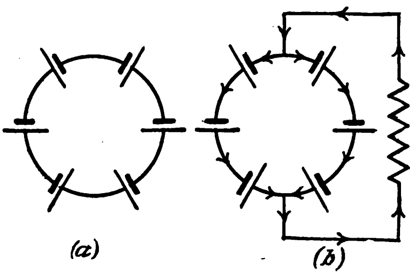


FIG. 281. — Batteries (*a*) without, and (*b*) with an external circuit.

But if we imagine the copper wires on the outer surface of the ring to be scraped bare, and if two metal or carbon blocks or **brushes** at the top and bottom rub on the wires as they pass, a current could be led out of the armature at one brush, and, after passing through an external resistance, such as a lamp, could be led back to the armature again at the other brush. In this case the armature circuit is *double*, consisting of its two halves in parallel. It is like adding an external circuit to the arrangement of cells described above. This battery analogue for a Gramme ring armature is shown in figure 281 (*b*).

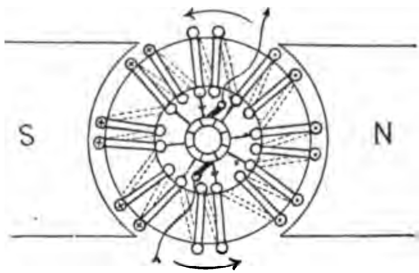


Fig. 282. — Ring armature with commutator.

In the Gramme ring arrangement there are at every instant the same number of active conductors in each half of the armature circuit, and so the current delivered by the armature is not only *direct* but also *steady*.

In practice, however, it would be difficult to make a good contact directly with the wires of the armature, because the wires must be carefully insulated from each other and from the iron core, and so the various turns of wire, or groups of turns, have branch wires which lead off to the commutator segments, as in figure 282.

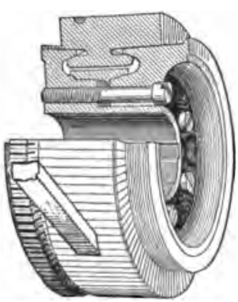


Fig. 283. — Commutator and brush.

The commutator consists of copper bars or segments which are arranged around the shaft and insulated from each other by thin plates of mica (Fig. 283). To get a satisfactorily steady current there should be many segments in a commutator, so that the brushes

may always be connected to the armature circuit in the most favorable way.

322. Drum armature. Since very little flux passes across the air space in the center of a Gramme-ring armature, the wires on the inner surface of the ring do not cut lines of magnetic force and are useless, except to connect the adjoining wires on the outer surface. Furthermore, it is very inconvenient to wind the wire on an armature of the ring form. For these reasons, most armatures are now of the **drum** type. In this form, the core is made with slots along the circumference for the wires to lie in (Fig. 284). Since the active wires in one slot are connected across the end to active wires in another slot, there are no idle wires inside the core.

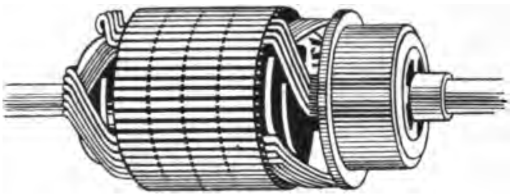


FIG. 284.—Slotted armature core, drum type.

323. Multipolar generators. The machines which have been described are called **bipolar** machines. For commercial purposes, especially in large machines, it is common practice to use four, six, eight, or even more poles. Such machines are called **multipolar.** By increasing the number of poles we can get the commercial voltages (110, 220, or 500 volts), at much slower speeds than would be necessary in a bipolar machine. We have already seen that the voltage depends on the rate at which the wires of the armature cut the lines of magnetic force. But in a four-pole machine (Fig. 285) each wire on the armature cuts a complete set of lines of

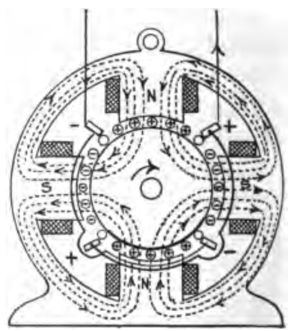


FIG. 285.—Four-pole generator.

force four times in each revolution instead of twice as in a two-pole machine. For this reason the speed of a four-pole machine is one half the speed required in a two-pole machine for the same voltage. Furthermore, the multipolar machine is more economical to build because it requires less iron to carry the magnetic flux. It will be observed from the diagram (Fig. 285) that every other brush is positive and is connected to the positive terminal of the machine.

QUESTIONS

1. If a person stands facing in the direction of the magnetic flux, and thrusts downward a wire which he holds in his two hands, in which direction is the induced e. m. f.?

2. What are the three factors which determine the voltage of a dynamo? How does each affect the voltage?

3. How many revolutions per minute (r. p. m.) would a single-coil bipolar dynamo have to make in order that the current might have 120 alternations per second?

4. How many revolutions per minute would an eight-pole generator have to make to have the current alternate 120 times a second?

5. Why are carbon blocks generally used instead of copper brushes?

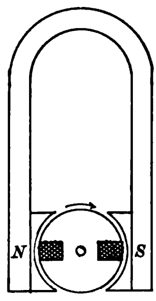


FIG. 286. — Magneto has permanent steel magnets.

324. Excitation of the field of generators. In the magneto (Fig. 286) the magnetic field is supplied by permanent steel magnets. In most other generators, the magnetic field is furnished by powerful electromagnets. Sometimes the current needed to excite these magnets is supplied by some outside source, such as a storage battery, but generally the machine itself furnishes the exciting current. There are three types of generators differing in the method of exciting the field coils; (1) *series-wound*, in which the whole current generated passes through the field coils on its way to

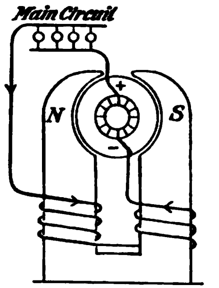


Fig. 287.—Series-wound generator supplying arc lamps.

the external circuit; (2) ***shunt-wound***, in which the field is excited by diverting a small part of the main current, the field coils and the external circuit being in parallel or in shunt; and (3) ***compound-wound***, which has both a series coil and a shunt coil.

In the **series** generator (Fig. 287), the field coils are wound with **a few turns** of **large** wire. When the current in the external circuit increases, the field is more highly magnetized, and so a higher voltage is available to supply the current. This machine is used to furnish current for arc lamps, which operate on a **constant current.**

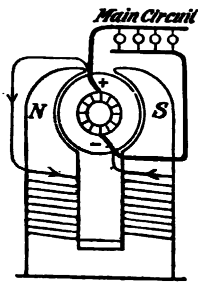


Fig. 288.—Shunt-wound generator supplying incandescent lamps.

When the field is **shunt-wound** (Fig. 288), the coils have **many turns** of **small** wire, for in this case it is desirable to divert as little current as possible from the main circuit, and so the resistance of the field coils should be high. Such machines are run at constant speed. When more load is thrown on the machine, that is, when more lamps are turned on, so that more current is needed, the terminal voltage drops a little. This decreases the current in the field coils and still further reduces the terminal voltage. A shunt machine, therefore, cannot be used when very constant voltage is desired.

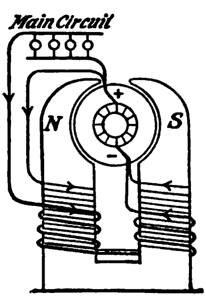


Fig. 289.—Compound-wound generator.

This drop in the terminal voltage of shunt generators under heavy loads can be overcome by the use of the **compound-wound** generator (Fig. 289), which is the one most

commonly used. Here the **voltage** is kept **constant** by adding a series coil of a few turns, which tends to raise the voltage when the current increases, just as in a series generator. If the coils are carefully adjusted, the voltage remains practically constant at all loads.

325. Source of energy in the dynamo. It is important to remember that the electric generator or dynamo can *not* of itself *make electricity*, but can only *transform* mechanical energy into electrical energy. For example, if we want to light a house with electricity, it is not enough for us to buy a dynamo, we must get also a steam engine, or a gas engine or a water wheel to *drive* the dynamo. We have already seen that the induced current is always in such a direction as to oppose the motion of the wire. Consequently, the greater the current in the dynamo, the greater the power needed to turn it. Large generators, such as are used in power stations to furnish electricity for street railways, sometimes require steam engines of 16,000 to 20,000 H. P. capacity.

Electric Power

326. How electric power is measured. To measure water power, we must know the quantity of water flowing per minute and the "head" of the water. Thus

Water power = quantity of water per minute × head.

$$\text{H. P.} = \frac{\text{lb. per min.} \times \text{ft.}}{33000}.$$

To measure electric power, we must multiply the quantity of electricity flowing per second — that is, the intensity of the electric current — by the voltage. Thus

Electric power = intensity of current × voltage.

The **watt** is the unit of electric power and may be defined as

the power required to keep a current of **one ampere** flowing under a drop or "head" of **one volt.**

Watts = amperes × volts.

Since the watt is a very small unit of power, we commonly use the **kilowatt** (K.W.) which is 1000 watts.

$$\textbf{K. W.} = \frac{\textbf{amperes} \times \textbf{volts}}{\textbf{1000}}.$$

Inasmuch as mechanical power is reckoned in horse power (H. P.), it will be convenient to know the relation of the unit of mechanical power to the unit of electrical power. Experiment shows that

1 horse power = 746 watts.
1 Kilowatt (K.W.) = 1.34 horse power (H. P.).

Since we have to compute electrical power very often, we may find a formula convenient.

$$P = IE,$$

where $P =$ power in watts,
$I =$ current in amperes,
$E =$ e. m. f. in volts.

For example, if a lamp draws 0.5 amperes from a 110 volt circuit, it is using power at the rate of 0.5 times 110 or 55 watts.

Again, suppose a street-car heater has a resistance of 110 ohms. At what rate is it consuming electricity on a 550 volt line? The current is $\frac{550}{110}$ or 5 amperes, and the power is 5 times 550 or 2750 watts or 2.75 K. W.

327. Commercial units of electrical work. Power means the rate of doing work. The total **work done** is equal to the **product** of the **rate** of doing work by the **time.** Thus if a steam engine is working at the rate of 15 horse power for 8 hours, it does 8 times 15 or 120 **horse-power hours** of work. In a similar way, if an electric generator is delivering electricity at the rate of 15 kilowatts for 8 hours, it does 8 times 15 or 120 **kilowatt hours** of work.

For example, we buy electricity by the kilowatt hour. In Boston the price is about 10 cents per kilowatt hour. If a store uses 100 lamps for 3 hours, each consuming electricity at the rate of 50 watts, it will cost

$$\frac{100 \times 3 \times 50 \times 0.10}{1000} = \$1.50.$$

328. Small units of electrical work. In the laboratory we often find it convenient to use a smaller unit of work, the **watt second** or **joule**.

Work (joules) = current (amperes) × e. m. f. (volts) × time (seconds).

Or $$W = IEt,$$

since 1 Kilowatt-hour = 3,600,000 watt seconds or joules,

1 Horse-power hour = 1,980,000 foot pounds.

Therefore **1 joule = 0.74 foot-pounds.**

PROBLEMS

1. How much electrical power (watts) is required to light a room with 5 lamps, if each lamp draws 0.4 amperes from a 110-volt line?
2. A street railway generator is delivering current to a trolley line at the rate of 1500 amperes and at 550 volts. At what rate (kilowatts) is it furnishing power?
3. How many horse power will be required to drive the generator in problem 2, if its efficiency is 90%?
4. A 10 kilowatt generator is working at full load. If the voltmeter reads 115 volts, how much does the ammeter read?
5. How many lamps, each of 120 ohms and requiring 1.1 amperes, can be lighted by a 25 K. W. generator?
6. How much power is required by a laundry using 5 electric flat-irons of 50 ohms each on a 110-volt line?
7. How much will it cost at 10 cents per kilowatt hour, to run a 220-volt motor for 10 hours, if the motor draws 25 amperes?
8. Would it be cheaper to buy the power needed in problem 7 at 8 cents per horse-power hour?
9. How much energy is consumed in a line whose resistance is 0.5 ohms, and which carries a current of 150 amperes for 10 hours?
10. How many joules of energy are consumed when a 40-watt lamp burns 10 minutes?

Electric Heating

329. Heating by electricity. We are familiar with the fact that electric cars are heated by electricity, and that an electric light bulb gets hot, and we may have used or seen electric flatirons (Fig. 290), electric ovens, or electric furnaces; but perhaps we do not realize that every electric current, however small, generates heat. This is because heat is generated so slowly in an electric bell, telegraph, or telephone, that it is radiated off without raising the temperature of the wires appreciably. It is this heating effect which limits the output of a generator, for if too heavy a current is drawn from the machine, the armature and field coils get so hot that the insulation is set on fire.

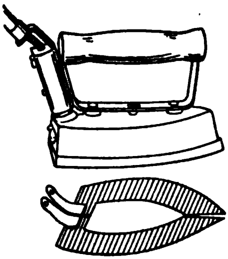


Fig. 290. — Flatiron heated by electricity, and the resistance wire in it shown separately.

330. Fuses and circuit breakers. To protect electrical machines from too much heat caused by excessive current, some sort of "electrical safety valve" has to be inserted in the circuit. Fuses are used for the small currents in house lights and small motors, and circuit breakers for larger currents in power stations. The essential part of a fuse is a strip of an alloy [Fig. 291 (*a*)], which melts at such a low temperature that the melted metal can do no harm. The size of the fuse

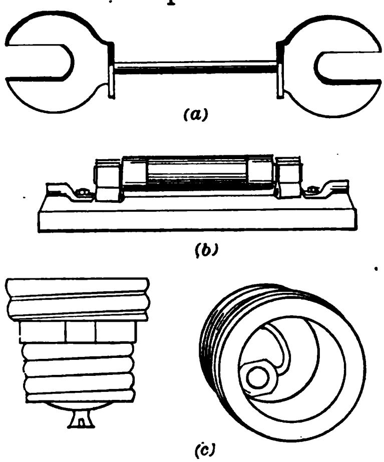


Fig. 291. — Fuses: (*a*) wire fuse; (*b*) cartridge fuse; (*c*) plug fuse.

is such that if by accident too heavy a current is sent through the wires, the fuse melts and breaks the circuit. At the moment the fuse melts there is an arc across the gap which might set things on fire. So the fuse is commonly inclosed in an asbestos tube, as in the "cartridge fuse" [Fig. 291 (*b*)], or in a porcelain cup, which screws into a socket like a lamp, as in the "plug fuse" [Fig. 291 (*c*)]. When the fuse wire melts because of excessive current, the fuse is said to "blow out."

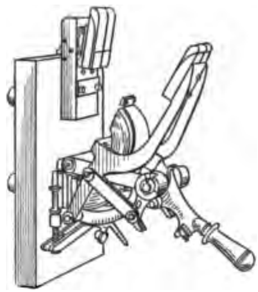


Fig. 292. — Circuit breaker.

A **circuit breaker** (Fig. 292) is simply a large switch which is automatically opened by an electromagnet when the current is excessive.

331. How much heat is generated by an electric current? The energy delivered to an electric heating coil, such as a flatiron, or soldering iron, is, as we have seen in section 326, EI joules per second, or EIt joules in t seconds. But since Ohm's law tells us that $E = IR$, if there is no cell, generator, or motor in the part of the circuit considered, we have the alternative statement, which is often more convenient in discussing electric heaters : —

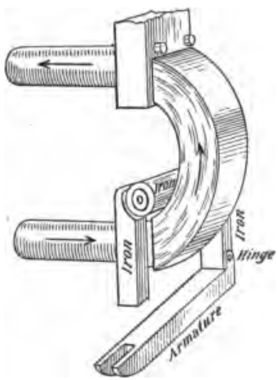


Fig. 293. — The electromagnet in the circuit-breaker in Figure 292. This type is used for very large currents, and needs only one "turn" of wire on the electromagnet.

Energy turned into heat $= I^2Rt$ (joules).

To measure the heat generated by an electric current in a wire, we can let it raise the temperature of a known weight of water. Careful experiments show that a current of

one ampere flowing through a wire of one ohm resistance for one second will generate enough heat to raise the temperature of one gram of water 0.24° C. That is,

$$H = 0.24\, I^2Rt,$$

where $H =$ heat in calories,
$I =$ current in amperes,
$R =$ resistance in ohms,
$t =$ time in seconds.

Problems

1. How many calories of heat are generated per hour in a 30 ohm electric flatiron using 4 amperes?

2. What is the cost of each calorie in problem 1, if the electricity costs 10 cents per kilowatt hour?

3. How much energy is turned into heat each hour by a current of 35 amperes in a wire of resistance 2 ohms? What size rubber-covered copper wire should be used to carry this current safely? (See table at end of Chapter XVI.)

4. If 88 % of the energy received by an electric lamp is converted into heat, how many calories are developed in one hour by a 35 candle power lamp drawing 0.9 amperes at 115 volts?

5. A 10 ohm coil of wire is used to heat 1000 grams of water from 15° C to 75° C in 10 minutes. How much current must be used?

6. How long will it take 5 amperes at 55 volts to raise the temperature of a kilogram of water from 20° to 100° C?

Electric Lighting

332. The electric arc. About a hundred years ago Sir Humphrey Davy (1778–1829), by using a battery of 2000 cells, made an electric arc between rods of charcoal. This was merely a brilliant lecture experiment, and it was not until sixty years later, when practical dynamos had been built, that arc lights became commercially possible. It was soon found that the coke which is formed in the ovens of gas furnaces makes a more durable material for the carbon than wood charcoal.

To show the form of the electric arc we may connect a current of 50 or more volts to two carbons, in series with a suitable rheostat. The light is so intense that the eyes must be shielded by blue glass from the direct glare. The arc can be projected on a screen with a convex lens. If D. C. current is used, the crater formed on the positive carbon and the cone on the negative carbon can be seen as shown in Fig. 294. The great heat evolved is shown by the fact that iron wire can be melted in the arc.

FIG. 294.— Positive and negative carbons of the arc.

Furnaces built on the principle of the electric arc are used to prepare artificial graphite, carborundum, calcium carbide, and various kinds of steel.

333. Modern arc lamps. Even coke carbon burns away, and so automatic lamps have been invented which feed their carbons gradually toward each other. Some of the early forms of these lamps made use of clockwork to feed the carbons, but now it is common to use a clutch which is worked by an electromagnet. One form of this mechanism consists of a "ballasting" resistance *B* (Fig. 295), which opposes any increase or decrease of current between the carbon tips, and of a "regulating" coil *S*, to control the distance between the carbon tips. When there is no current, the plunger *P* drops and releases the friction clutch on the upper carbon *C*. When the current is on, the plunger *P* is pulled up and lifts the clutch and upper carbon the proper distance.

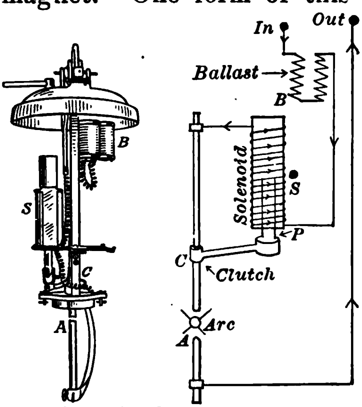


FIG. 295. — Arc lamp, and diagram of automatic feed.

Recently an **inclosed arc** lamp (Fig. 296) has come into general use. When the arc is surrounded by a glass globe which is nearly air-tight, the available supply of oxygen is quickly used up and the same pair of carbons lasts 100 hours instead of only 7 or 8 hours.

FIG. 296.—Inclosed arc lamp.

If the carbon rods are made with a core of calcium fluoride, the vapor given off is very luminous and gives a light of a golden orange color. These so-called **flaming arcs** are extensively used on streets for advertising purposes. In this type the carbons are long and slender, and both carbons feed down, as shown in figure 297.

FIG. 297.—Impregnated carbons of flaming arc.

Another form of flaming arc is the **magnetite arc.** In this lamp the lower electrode is made of magnetite or some similar substance, powdered and compressed in a sheet-iron tube, while the upper electrode is of solid copper which wastes away very little.

In the **mercury arc** or Cooper-Hewitt lamp, use is made of the luminescence of mercury vapor. The mercury is held in the lower end of a glass vacuum tube 2 to 4 feet long (Fig. 298). Some special device has to be used to start the current through the mercury vapor; but once started, the current flows easily through the hot vapor, which glows with a light composed of green, blue, and yellow, but *no red.*

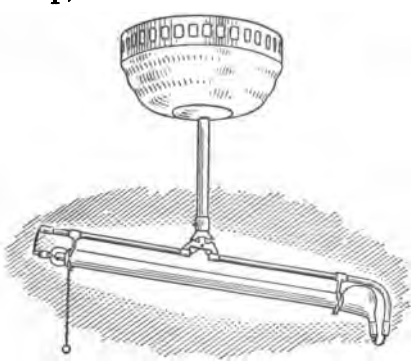


FIG. 298.—Mercury arc lamp, started by tilting.

This gives a peculiar color to objects thus illuminated. (See Chapter XXIII.)

334. Carbon filament lamps. In **incandescent** lamps, there is a wire or **filament** which is heated white hot by the electric current. The light emitted is the same as it would be if the same wire could be heated to the same temperature in any other way, as by an oxyhydrogen flame. In the early experiments platinum was tried for the filament, but even though its melting point is as high as 1600° C, it could not stand the temperatures required. Carbon is one of the few known substances with a higher melting point, and in 1880 Edison and others succeeded in making a lamp with a carbon filament. Since the filament would burn up at once if there were any air present to support the combustion, it has to be inclosed in a glass bulb (Fig. 299) in which there is a very high vacuum.

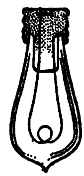


FIG. 299.—Carbon filament lamp.

The electricity is led into and out of the filament through two short platinum wires, melted into the glass bulb at one end. These platinum wires are connected by copper wires to the brass collar and metal tip at the end of the bulb. Such lamps are usually made for 110 or 220 volts. If a lamp made for 110 volts is used on a 105 volt line, it will probably last twice as long, but will only give 80% as much light. If it is used on a 113 volt line, even though it gives about 18% more light, it will last only half as long. So it is very desirable to use lamps on the voltage for which they are intended. This means we must have *good regulation* on the *electric lighting service;* that is, constant voltage at all loads.

335. Commercial rating of electric lights. To measure the **output** of light from a lamp, we need some standard lamp for comparison. As will be explained in Chapter XXI, this standard is the so-called **international candle.** The ordinary

incandescent light is equivalent to about 16 such standard candles, and it said to be 16 candle power (16 c.p.). Since lamps do not show the same brightness on all sides, it is customary to take the average candle power in all directions in a horizontal plane. Thus incandescent lamps are often rated according to their **mean horizontal candle power.**

The **input** of electrical energy is measured in watts. The **commercial rating,** which is also called the "*efficiency*,"* of electric lamps is the number of **watts per candle power.** For example, an ordinary 50 watt lamp gives 16 candle power; so its commercial rating is $\frac{50}{16}$ or 3.1 watts per candle power.

By a special firing process, carbon filaments can be "metallized" The efficiency of such filaments is about 2.5 watts per candle power.

336. Metal filament lamps. Still more efficient incandescent lamps are made with **metallic** filaments. The metals most used are *tantalum* and *tungsten*, both of which have melting points much higher than that of platinum. Since their specific resistance is much lower than that of carbon, a metal filament must be very much longer and thinner (about 0.02 millimeters in diameter) than a carbon filament to have the necessary resistance. So long a wire can be put in a bulb of the ordinary size only by winding it zigzag on star-shaped reels, as shown in figure 300.

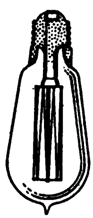


FIG. 300. — Metal filament lamp.

One difficulty with these metal filaments is their brittleness and liability to breakage. Furthermore, they soften somewhat when hot, and if a metal-filament lamp is used in a horizontal position, the filament may sag and short-circuit. Nevertheless, the extremely high efficiency of these lamps, their long life (except for breakage), and their wonderful

* This is really a measure of inefficiency; the larger the number the worse the lamp.

white light, which is the same color as daylight, have made them very popular.

COMPARATIVE "EFFICIENCY" OF ELECTRIC LAMPS

NAME OF LAMP	WATTS PER CANDLE POWER	NAME OF LAMP	WATTS PER CANDLE POWER
Carbon filament . .	3 to 4	Arc lamp	0.5 to 0.8
Metallized carbon . .	2.5	Mercury arc	0.6
Tantalum	2.0	Flaming arc	0.4
Tungsten	1.0 to 1.5		

QUESTIONS AND PROBLEMS

1. Why must a rheostat be used in series with the arc lamp in a projection lantern?

2. Why are the flaming arc lamps, which are used for street lighting, placed high above the street?

3. When an incandescent light bulb gets very hot and blackens on the inside, what does it indicate?

4. What must be the voltage of an arc-lighting dynamo which is to furnish 8 amperes to 25 street lamps arranged in series, if each lamp requires a terminal voltage of 50 volts?

5. What would be the kilowatt output of the generator in problem 4?

6. How may a street car, which is operated on a 550 volt line, be lighted by 110 volt lamps? Draw a diagram of the connections.

7. In considering the proper kind of electric lamp for illumination, what other factors must be considered besides watts per candle power?

8. How many 0.5 ampere lamps, connected in parallel, can be protected by a 20-ampere fuse?

9. How many candle power should a 50 watt tungsten lamp give, if its efficiency is 1.2 watts per candle power?

10. It was found on testing a 32 candle power lamp that it consumed 100 watts of electric power, of which 88 watts were turned into heat. What was its efficiency for heating? What was its light efficiency in the true sense? What was its commercial rating?

Electric Motor

337. The dynamo as a motor. We have already seen that a dynamo, when driven by a steam engine, gas engine, or water wheel, may *generate* electricity. Now we shall see how this electric current can be supplied to a second machine, exactly like a dynamo, but called a **motor**, which may be used to drive an electric car, a printing press, a sewing machine, or any other machine requiring mechanical energy. In short, the dynamo is a *reversible* machine, and sometimes in shops, and often on self-starting automobiles, the same machine is driven as a generator part of the time, and used as a motor to drive another machine the rest of the time.

Structurally, the motor, like the dynamo, consists of an electromagnet, an armature, and a commutator with its brushes. To understand how these act in the motor, however, we must get a clear idea of the behavior of a wire carrying an electric current in a magnetic field.

338. Side push of a magnetic field on a wire carrying a current. We will stretch a flexible conductor loosely between two binding

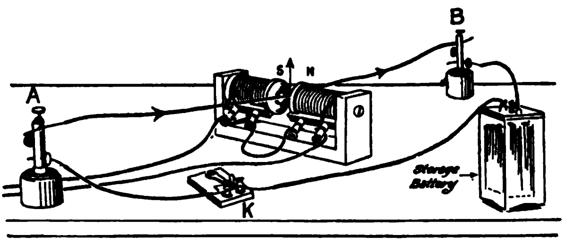


FIG. 301. — Side push on wire carrying a current.

posts A and B, so that a section of the conductor lies between the poles of an electromagnet, as shown in figure 301. Let the exciting current be so connected to the electromagnet that the poles are N and S as shown. Then, if a strong current from a storage battery is sent through the conductor from A to B by closing the key K, it will be seen that the wire

between the magnet's poles is instantly thrown *upward*. If the current is sent from B to A, the motion of the conductor is reversed, and it is thrown *downward*.

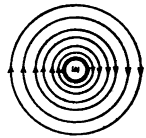


FIG. 302. — Magnetic field about a wire with current going in.

It will help us to understand this side push exerted on a current-carrying wire in a magnetic field, if we recall that every current generates a magnetic field of its own, the lines of which are concentric circles. Figure 302 shows a wire carrying a current *in*, that is, at right angles to the paper and away from us. The lines of force are going around the wire in clockwise direction.

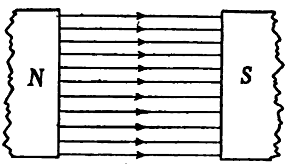


FIG. 303. — Uniform magnetic field.

The magnetic field between the poles of a strong magnet is practically uniform and is represented by parallel lines of force shown in figure 303.

If we put the wire, with its circular field, in the uniform field between the N and S poles of the magnet, the lines of force are very much more crowded above the wire (Fig. 304) than below. But we have seen in section 238 that we can think of magnetic lines of force as acting like rubber bands which would, in this case, push the wire down. If the current in the wire is reversed, the crowding of the lines of force comes below the wire, and it is pushed up.

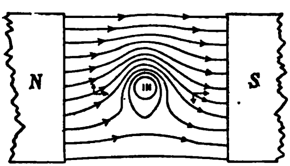


FIG. 304. — Lines of force about a wire carrying current in a magnetic field.

339. Motor rule of three fingers. The rule for remembering which way this side push on a wire in a magnetic field will move the wire is precisely the same as that for the generator except that the **left hand** instead of the right is used.

340. The action of a motor. In motors, as in dynamos, the drum type of armature is almost exclusively used. It will be remembered (see section 322) that in this type the active wires lie in slots along the outside of the drum, as in figure 305, and the wiring connections across the ends of the armature are such that when the current is coming *out* on one side, — say the right, — it will be going *in* on the other side — the left. Just how these wiring connections are made is not important for the present purpose, and indeed there are many different ways in which they can be arranged. In any case, from what has just been said, it will be clear that the wires (⊙) on the right side of the armature will be pushed *upward*, and those (⊗) on the left side of the armature will be pushed *downward* by the magnetic field. In other words, there will be a **torque** tending to rotate the armature counter-clockwise. The amount of this torque depends on the number and length of the active wires on the armature, on the current in the armature, and on the strength of the magnetic field.

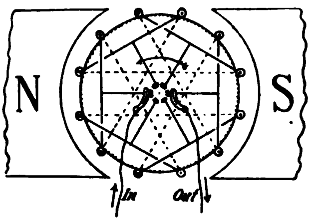


Fig. 305. — Drum-wound motor.

Another way of looking at this action is to notice that the effect of these armature currents is such as to make the armature core a magnet, with its north pole at the bottom and its south pole at the top. The attractions and repulsions between these poles and those of the field magnet cause the armature to rotate as indicated by the arrows.

The function of the commutator and brushes is, as in the generator, to reverse the current in certain coils as the armature rotates, so as to keep the current circulating, as shown in figure 305.

341. Forms of motors. D.C. generators and motors are often of identical construction. Thus we have **series motors,**

such as are used on street cars and automobiles, and **shunt** motors, such as are used to drive machinery in shops. So also we have **bipolar** and **multipolar** motors. When it is desirable that a motor shall run at a slow speed, it is built with a large number of poles.

342. Back e. m. f. in a motor. Suppose we connect an incandescent lamp in series with a small motor. If we hold the armature stationary, and throw on the current from a battery, the lamp will glow with full brilliancy, but when the armature is running, the lamp grows dim.

This shows that *a motor uses less current when running than when the armature is held fast.* The electromotive force of the battery and the resistance of the circuit are not changed by running the motor. Therefore, the current must be diminished by the development of a **back electromotive force**, which acts against the driving e. m. f.

Since a motor has a series of armature wires cutting magnetic lines of force, it is bound to generate an e.m. f. in these wires. That is, every motor is at the same time a dynamo. The direction of this induced e. m. f. will always be opposite to that driving the current through the motor.

Just as in the generator, when the armature revolves faster, the back e. m. f. is greater, and the difference between the impressed e. m. f. and the back e. m. f. is therefore smaller. This *difference* is what drives the current through the resistance of the armature. So a motor will draw more current when running slowly than when running fast, and much more when starting than when up to speed.

For example, suppose the impressed or line voltage on a motor is 110 volts, and the back e.m. f. is 105 volts. Then the *net* voltage which will force current through the armature is 110 – 105, or 5 volts. If the armature resistance is 0.50 ohms, the armature current is 5.0/0.5, or 10 amperes. But if the whole voltage (110 volts) were thrown on the armature while at rest, the current would be 110/0.5 or 220 amperes.

343. Starting a motor. When a motor starts from rest, there is, of course, no back e. m. f. at first, and if the motor

is thrown directly on the line, there will be such an excessive current as to "burn out" the armature. To prevent this first rush of current, a *starting resistance* is put into the circuit at first, and cut out step by step as the machine speeds up. The device for doing this is shown in figure 306. See also figure 257.

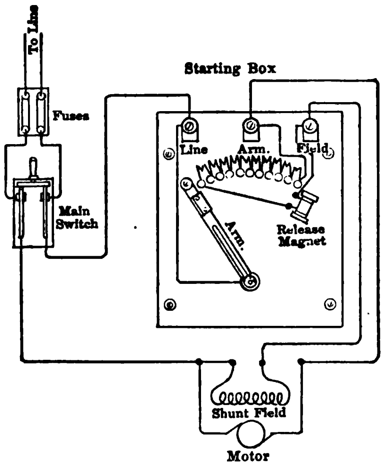


FIG. 306.—Motor with starting resistance.

344. Applications of the motor. The transmission of power through shops and factories by means of shafting, cables, and belts is dangerous, noisy, and uneconomical. In a modern system, electric power is generated in a central power house, is transmitted to various parts of the plant, and is used in electric motors to drive either individual machines or groups of machines. When electrical transmission is used, the danger and inconvenience of belts and shafting are avoided, the machines can be set in any position, and their speed can be easily controlled by field rheostats. In shops and factories thus equipped, **shunt** motors are commonly used, for constant speed motors are required, and the speed of a shunt motor under no load, or a light load, is nearly the same as at full load.

Series motors are used on cranes, automobiles, and electric cars, because this type of motor has a **large starting torque.** The torque in a series motor is proportional to the *square* of the current, while in a shunt motor it is *directly* proportional to the current. The fact that the torque in a series motor is largest when the speed is slowest (because there is

little back e. m. f.) makes it just the kind of motor for crane or vehicle work. When the load on a series motor drops to zero, the motor may "race"; that is, go faster and faster until the armature flies to pieces. For this reason, series motors are connected, either directly (on same shaft) or by cogwheels, to the machines to be driven, so that they can never escape their load.

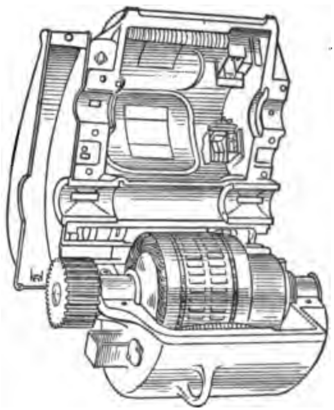


Fig. 307. – Street-car motor with top of case lifted up.

Figure 307 shows a street-car motor with its case lifted to show the inside arrangement. The field consists of *four* short poles projecting from the case, which serves both to protect the motor, and as a path for the magnetic flux. The armature revolves so rapidly that its speed has to be reduced by a pair of cogwheels, the larger of which is on the axle of the driving wheels, and is not shown in the picture. These make the speed of the axle about one fourth that of the motor.

Street cars are usually operated on a direct-current system. A large multipolar compound-wound generator (Fig. 308) at the power station maintains about 550 volts between the trolley or third rail and the track. A "feeder" or cable of low resistance is run parallel to the trolley wire and connected to it at intervals, to avoid a large voltage drop in the line when a number of cars are taking current at a distance from the power plant. The current passes down the trolley pole

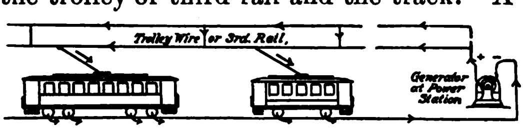


Fig. 308. — General scheme of a trolley line.

into the **controller** (Fig. 309). This is an ingenious arrangement of switches by which the motorman can start his car with both motors in series and with the starting resistance all in; then by moving a lever he gradually cuts out the starting resistance and finally throws both the motors in parallel, as shown in figure 310. Thus, when starting, each motor receives less than half the line voltage, and when running at full power, gets full voltage. The current leaves the motors by the wheels, and goes back to the power station through the rails.

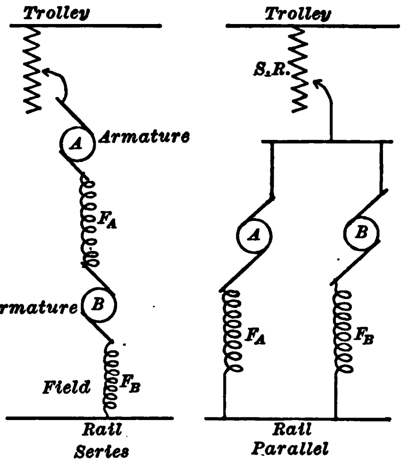


FIG. 310.—Series-parallel control of electric cars.

345. Efficiency of the electric motor. One reason for the extensive use of electric motors is their great efficiency, sometimes as high as 80% or 90%. The **efficiency** of a motor, just as of any machine, means the *ratio of out-put to in-put.* We can easily measure the number of amperes and the number of volts supplied to the motor and thus compute the **watts put in.**

To get the **output** of mechanical work, engineers usually make a "brake-test." One simple form of brake consists of a belt or cord attached to two spring balances and passing under a pulley on the motor shaft, as shown in figure 311.

If the pulley rotates as indicated, it is evident that one spring balance will have to exert more force than the other because of the friction of the pulley on the cord. *The amount*

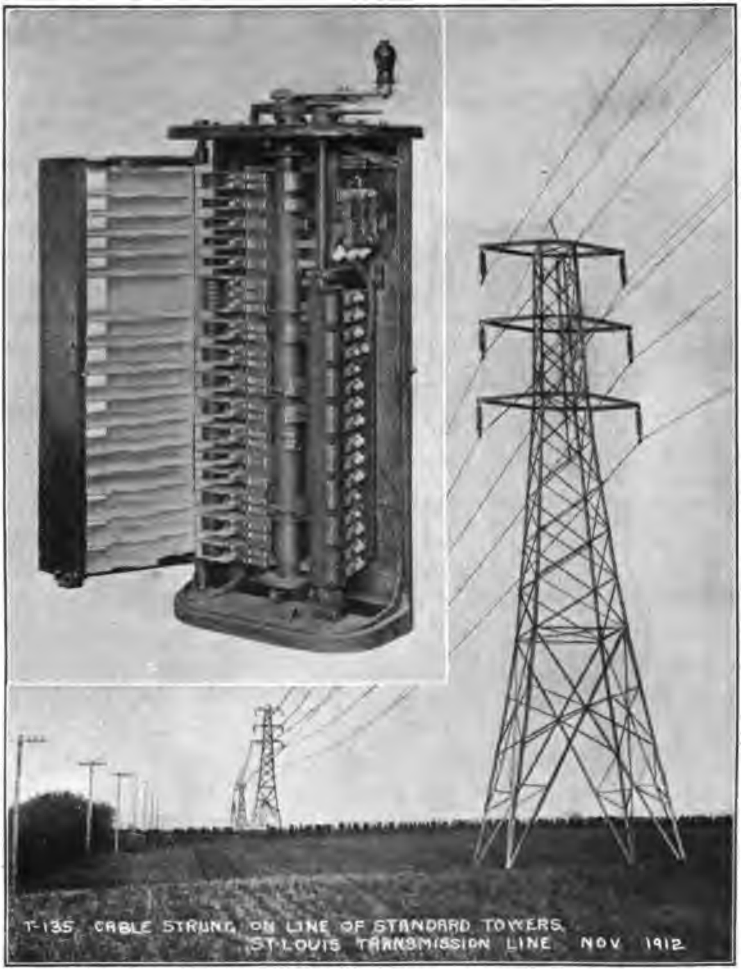
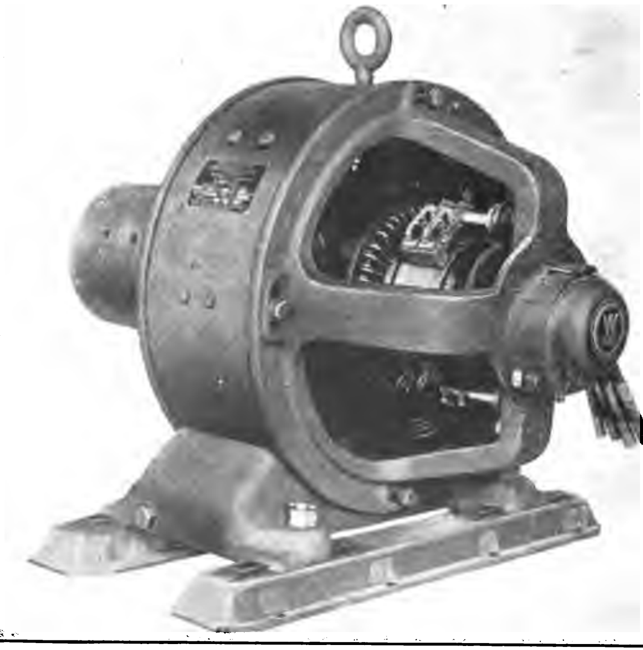


Fig. 309 (above at left). — Street car controller.

Fig. 324 (below at right). — Transmission line.



Direct Current Motor with field coils shown above.

of friction is equal to the difference between the readings of the two balances, and it is exerted each minute through a distance equal to the circumference of the pulley times the revolutions per minute. The **work done** in one minute is equal to the friction times the distance per minute.

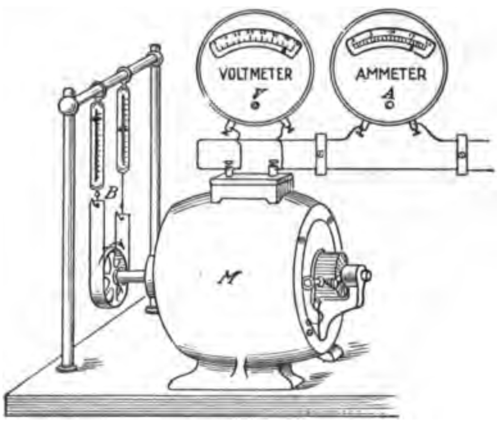


FIG. 311. — Measuring output of a motor by means of a brake.

Finally, if we express the output and input in some common unit of power and divide, we have the efficiency. It will be helpful to know that

1 watt = 44.3 foot pounds per min. = 6.12 kilogram meters per min.

QUESTIONS AND PROBLEMS

1. Figure 312 represents a bipolar motor with the armature revolving counter-clockwise. Copy it and indicate by dots and crosses* in circles, the direction of the various currents.

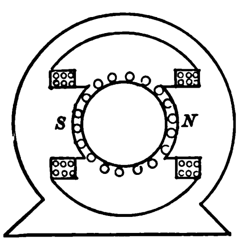


FIG. 312 — Bipolar motor.

2. What is the armature resistance of a motor in which the armature current is 4 amperes, the impressed e. m. f. is 115 volts, and the back e. m. f. is 112 volts?

3. Find the back e. m. f. in a motor in which the armature resistance is 0.3 ohms, the current is 15 amperes, and the impressed voltage is 110 volts.

* A cross in a circle represents the feathers of an arrow sticking into the paper, and means current going *in*. A dot in a circle means a current coming out.

4. How much current will be drawn by a motor whose efficiency is 90%, when it is developing 5 H. P. and is connected to the 110 volt service?

5. When a certain motor was tested by the brake test, it took 67 amperes at 113 volts and developed 8.5 H. P. Calculate its efficiency.

Chemical Effects of Electric Currents

346. Conduction by solutions. When an electric current flows along a copper wire, the wire becomes warm and is surrounded by a magnetic field. When an electric current flows through a solution of salt and water, the solution is warmed and is surrounded by the magnetic field, *and* it is at the same time **decomposed** or broken up. For example, under certain conditions an electric current will decompose brine into a metal, sodium, and a gas, chlorine, which are the two elements composing salt. Not all liquids conduct electricity; thus alcohol and kerosene are nonconductors. But all liquids which do conduct electricity are more or less decomposed in the process.

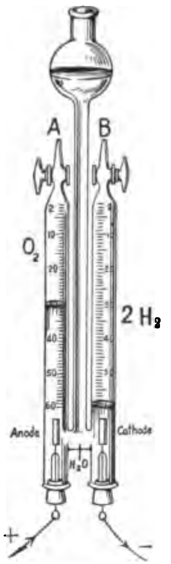


FIG. 313. — Water is broken up into oxygen and hydrogen.

347. Electrolysis of water. Water (made slightly acid with sulphuric acid) can be decomposed by an electric current in the apparatus shown in figure 313. The platinum electrodes are connected with a battery or generator, giving at least 5 or 6 volts. The electrode in tube A, which is connected to the positive (+) pole, is called the *anode*, and the other electrode in B is the *cathode*. The current passes through the solution from the anode A to the cathode B. Small bubbles of gas are seen to rise from both electrodes, and the gas collects in tube B twice as fast as in tube A. When tube B is full, we open the switch, and test the collected gases. To test the gas in tube B, we open the stopcock at the top and apply carefully a lighted match. This gas burns with a pale

blue flame which shows it to be *hydrogen*. If we open the stopcock in tube *A* and bring a glowing pine stick near, it bursts into a flame, which shows the gas to be *oxygen*.

Thus we see that water is decomposed by electricity into its constituent elements, hydrogen and oxygen. This process of decomposing a compound by means of an electric current is called **electrolysis**.

348. Theory of electrolysis. The theory of this process may be stated as follows: The small quantity of sulphuric acid (H_2SO_4), when put into the water, breaks up into hydrogen ions ($2\ H^+$) and sulphate ions (SO_4^{--}), which have positive and negative charges of electricity respectively. When the current is sent through the solution, the positive hydrogen ions ($2\ H^+$) wander toward the cathode and the negative sulphate ions (SO_4^{--}) toward the anode. At the cathode, the hydrogen ions give up their positive charges and rise to the surface as bubbles of hydrogen. At the anode, the sulphate ions give up their negative charges of electricity and react with the water (H_2O) to form sulphuric acid (H_2SO_4) and to set free oxygen (O_2). In this way the sulphuric acid, which is added to conduct the electricity, is not used up, while the water ($2\ H_2O$) is broken into hydrogen ($2\ H_2$) and oxygen (O_2).

349. Electroplating. We may illustrate the process of electroplating by the following experiment.

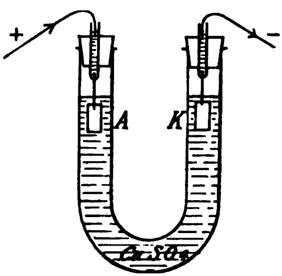


FIG. 314. — Electrolysis of copper sulphate.

We will put two platinum electrodes in a U-tube filled with copper sulphate solution ($CuSO_4$), as shown in figure 314. After the electric current has passed through the solution for a few minutes, we find the cathode is coated with metallic copper, while the anode is unchanged. If we reverse the direction of the current, we find that copper is deposited on the clean platinum plate which is now the cathode, and the copper coating on the anode gradually disappears.

In this way one metal can be coated with another. For example, articles of brass and iron, which corrode in the air, can be coated with nickel, which does not corrode. Similarly, much cheap jewelry is gold or silver plated. Many knives, forks, and spoons are silver plated, the best being what is called "triple" or "quadruple plate."

In practice the process is done in vats, as in figure 315. The objects to be plated are hung from one copper "bus" bar, and the metal to be deposited, in this case pure silver, is hung from the other bar. The vat contains a solution of the metal to be deposited. For silver plating a solution of silver and potassium cyanide is used. The bar carrying the metal to be deposited is connected with the + terminal of a low-voltage generator and the other bar to the − terminal. The silver plates at the anode dissolve as fast as the silver is deposited on the cathode, the strength of the solution remaining unchanged. When the coating has reached the proper thickness, a final process of buffing and polishing gives the surface the desired appearance.

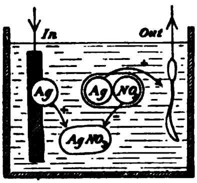


FIG. 315. — Diagram of electroplating vat.

350. Electrotyping. One might at first suppose that this book was printed from the actual type which was set up, but that is not the case. Most books which are made in large numbers are printed from electrotype "plates." A wax impression of the page as set up in type is made in such a way that every letter leaves its imprint on the wax mold. Since the wax is itself a non-conductor, it has to be coated with graphite. This mold is then placed in a solution of copper sulphate and attached to the negative bus bar, so that it becomes the cathode, while a copper plate acts as the anode. After the current has deposited copper on the wax mold to the thickness of a visiting card, this shell of copper is separated from the mold and "backed up" with type metal to give it the necessary strength for printing.

351. Refining of metals. Copper as it comes from the smelting works is not pure enough for some purposes, such as making wires and cables for carrying electricity. So the copper for electrical machinery is refined by electricity. The crude copper is the anode, a thin sheet of pure copper is the cathode, and the solution is copper sulphate. The copper deposited by the electric current is remarkably pure. The anode of crude copper gradually dissolves, and the impurities drop to the bottom of the vat as mud. In this mud there is generally enough gold and silver to pay the expense of the process. Copper purified in this way is known commercially as **electrolytic copper**.

352. Electrochemical equivalents of metals. Experiments show that a given current always deposits the same amount of a given metal from a solution in a given time. In fact, this is so accurately true that it is the basis of the most accurate method known for calibrating standard ammeters (see section 275). The amount of metal deposited by a current depends (1) on the strength of the current, (2) on the time it flows, and (3) on the nature of the metal. The definite quantity of a substance deposited per hour by electrolysis when one ampere is flowing through a solution is called the **electrochemical equivalent** of the substance.

Electrochemical Equivalents

Element	Symbol	Grams per Ampere Hour
Aluminum	Al	0.337
Copper	Cu	1.186
Gold	Au	3.677
Hydrogen	H	0.0376
Nickel	Ni	1.094
Oxygen	O	0.298
Silver	Ag	4.025

Questions and Problems

1. To determine which is the + and which the − pole of a generator, two copper wires are sometimes connected to the terminals and the bared ends dipped in a glass of water. One will soon turn dark. How does this experiment show which is the positive terminal?

2. How many grams of silver are deposited in 8 hours from a silver nitrate solution by a current of 5 amperes?

3. How many liters of hydrogen will be generated by a current of 10 amperes in 4 hours? (A liter of hydrogen weighs 0.09 grams under standard conditions.)

4. How many amperes will be needed to deposit 1.5 pounds of copper per day of 24 hours?

5. How long will it take a current of 200 amperes to refine a ton of copper?

6. In calibrating an ammeter the current was allowed to run 2 hours and 15 minutes, and deposited 39.5 grams of silver. What would be the reading of the ammeter, if correct?

7. Two electroplating vats are arranged in series, one for gold and the other for silver. How much gold is deposited while 1 gram of silver is being deposited?

8. An electroplater buys his electricity by the kilowatt hour. The metal deposited in electroplating is proportional to the number of ampere hours. Why does he use as low a voltage as possible?

9. What is meant by triple and quadruple plate?

353. Storage battery. Some people think a storage battery is a sort of condenser where electricity is stored, but it is not that. In the storage battery, as in any other battery, the electrical energy comes from the chemical energy in the cells. The **charging** process consists in forming certain chemical substances by passing electricity through a solution, just as hydrogen and oxygen are formed in the electrolysis of water. In the **discharging** process, electricity is produced by the chemical action of the substances which have been formed in the charging process.

354. Lead storage cell. We may make a small lead storage cell by putting two sheets of ordinary lead in a glass battery jar with a very dilute solution of sulphuric acid. To charge it or "form" the plates

quickly we connect this cell and an ammeter in series with a battery of three or more cells, or better, a generator of about 6 volts (Fig. 316). While the current is passing, bubbles of gas rise from each plate. If, after a few minutes, we disconnect the generator and touch the wires of a voltmeter to the lead terminals, it shows an e. m. f. of about 2 volts. If we then connect an electric bell in series with the ammeter and the lead cell, the bell rings, which shows that a current is produced, and the ammeter shows that the current on discharge is opposite to that used in charging the cell. When the plates are lifted out of the solution after charging, plate *B*, the anode, is *brown*, due to a coating of lead peroxide (PbO_2), and plate *A*, the cathode, is the usual gray of pure lead (Pb).

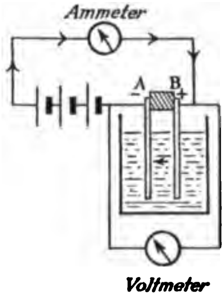


FIG. 316. — Forming a lead storage cell.

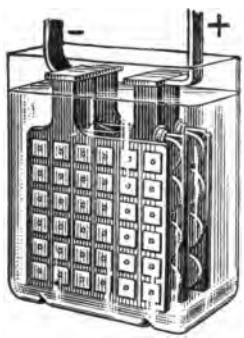


FIG. 317. — Commercial lead storage cell.

In the commercial lead storage (Fig. 317) cell, the negative plates are pure spongy lead (Pb), the positive are lead peroxide (PbO_2), and the electrolyte is dilute sulphuric acid. In the charging process, the positive plate, which is dark brown, is coated with lead peroxide, and the negative, which is gray, is made into spongy lead. In the discharging process, both plates gradually return to a condition where each is covered with lead sulphate ($PbSO_4$). This is shown in figure 318. The chemistry of these changes can be briefly described by the equation

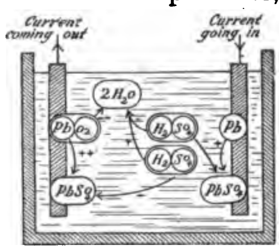


FIG. 318. — Discharging a lead cell.

$$\xleftarrow{\text{Charge}}$$
$$PbO_2 + Pb + 2\,H_2SO_4 = 2\,PbSO_4 + 2\,H_2O.$$
$$\xrightarrow{\text{Discharge}}$$

It will be noticed that during the charging process the acid becomes more concentrated. So the condition of a storage cell can be determined, at least roughly, by the specific gravity of the acid. The plates in the commercial lead battery are either roughened and then changed into the proper active materials, lead peroxide and lead, by a chemical process, or are punched full of holes which are filled with the active material.

355. Advantages and disadvantages of the storage cell. The lead cell is heavy and expensive, and requires careful handling to get an efficiency even as high as 75%. Its principal use is not as yet for automobiles, but in three other fields. First, it is often used to carry the "peak" of the load of a power station. In certain hours of the day the demand for current is too great for the generators to carry, so a large storage battery, which has been charging while the load was light, is used to help out the generators. Second, many companies, which have to furnish electricity without interruption or pay a heavy fine, use a storage battery as a reserve supply of electrical energy. In case of accident, the storage battery can be drawn upon at a moment's notice. Third, in some small plants the load on the generators is very light for a considerable time each day or night. In such cases a storage battery is sometimes used to take care of this long-continued light load, and the engine and generators are shut down.

356. Edison storage battery. Edison has invented a storage cell in which the negative plate is pure iron in a steel frame, the positive plate is nickel oxide, and the solution is caustic potash. Since this cell is intended for traction work, great pains have been taken to make it light, strong, and compact. Instead of being placed in a glass or hard-rubber tank, it has a thin nickel-plated sheet-steel case. In a lead cell the normal voltage on discharge is 2 volts; in the Edison cell it is 1.2 volts. For the same capacity of output,

the Edison cell is about half as heavy as the lead cell. As the internal resistance of the Edison is a little more than that of the lead cell, its efficiency is a little lower. Whether or not the Edison cell is going to be better than the lead cell depends on its "life" under commercial conditions, and this is not yet settled.

Questions and Problems

1. In a trolley system the generator maintains 565 volts on the line. How many lead storage cells, each of 2.1 volts, will be needed to help the generator carry the peak of the load?
2. Storage cells are sold according to their "capacity" in ampere hours. What "capacity" will be required to deliver 10 amperes continuously for 8 hours?
3. Most manufacturers of lead cells allow about 55 ampere hours for each square foot of positive plate area. How large a plate area will be required in problem 2?
4. If the e.m.f. of a lead cell is 2.3 volts on open circuit, while the terminal voltage when the cell is delivering 10 amperes is only 2 volts, what is the internal resistance of the cell?
5. A battery of 24 lead storage cells in series, each having an e.m.f. of 2.1 volts, a normal charging rate of 15 amperes, and an internal resistance of 0.005 ohms, is to be charged by a dynamo, what must be the terminal voltage of the dynamo?

SUMMARY OF PRINCIPLES IN CHAPTER XVIII

When a wire cuts lines of force, an induced e.m.f. is set up in the wire.

To get direction of current, use *right* hand.

Thumb = Motion,
Forefinger = Flux,
Center finger = Direction of Current.

Magnitude of e.m.f. varies as speed × flux × turns.

Slip rings give alternating current.
Commutator gives direct current.

Dynamo does not *make* energy; it *transforms* mechanical into electrical energy.
Motor transforms electrical energy into mechanical energy.

Power delivered to circuit = intensity of current × voltage.

Watts = amperes × volts.

1 H. P. = 746 watts.

Power turned into heat = current squared × resistance.

$$\text{Watts} = (\text{amperes})^2 \times \text{ohms}.$$
$$\text{Heat in calories} = 0.24\ I^2Rt.$$

A wire carrying a current, when set at right angles to a magnetic field, is pushed sideways by the field.

To get direction of motion, use *left* hand. As before,

Thumb = Motion,
Forefinger = Flux,
Center finger = Current.

Every motor, when running, is acting at the same time as a dynamo. The e.m.f. of this dynamo action opposes the current driving the motor, and is the *back* e.m.f.

Net e.m.f., which drives current through armature, equals impressed e.m.f. minus back e.m.f.

Ohm's law applies to a motor armature *only* if *net* e.m.f. is used.

Weight of a substance deposited by a current
= electrochemical equivalent × current × time.

Questions

1. Why cannot a lead storage cell be charged from a dry cell?
2. Why do the lights on an electric car often grow dim when the car is crowded and going up grade?
3. Would it be possible to drive the propellers of an ocean liner by electric motors? Why is it not commonly done? Why are some people seriously considering doing it in the near future?

4. Which will yield the more heat for warming an electric car, a 50 ohm resistance connected across a 50 volt line, or a 100 ohm resistance connected across a 100 volt line?

5. Compare the cost per hour of running a 55 ohm electric heater on a 55 volt circuit and on a 110 volt circuit, if power costs 10 cents per kilowatt hour.

6. The "carrying capacity" of a wire is limited by the rate at which it can radiate the heat generated in it. Which will require wires of larger carrying capacity, a 1100 volt power transmission line, carrying 1000 amperes, or a 11,000 volt line, carrying 100 amperes?

7. Which of the lines in the last problem will deliver more power at the other end?

8. Why are electric cars not more generally operated on storage cells instead of by an overhead or a third-rail system of transmission?

9. Why are electric light bills made out in kilowatt hours instead of kilowatts?

10. Why does it take twice as much power to keep a generator going when there are 200 incandescent lamps lighted in parallel as when there are only 100 lamps in use?

11. What methods are used to make the track of a street-car system a better conductor?

12. If you were to charge a storage battery so incased that you could see only the two terminals which were marked + and –, how would you connect it to a generator?

13. How does the back e. m. f. of a motor vary with its speed?

14. A belt-driven shunt dynamo is used to charge a storage battery. The belt breaks, but the dynamo keeps on running. Explain.

15. Does it make any difference which end of the field coils of a shunt-wound dynamo is connected with the positive brush? If you have an experimental dynamo, try it.

16. The speed of a shunt-wound motor can be controlled by putting an auxiliary resistance, called a field rheostat, in series with its field coils, so as to decrease the current through them. Will this increase or decrease its speed? Why? If you have an experimental motor, try it.

17. What is the advantage of electrotype plates over the original type in printing a book?

CHAPTER XIX

ALTERNATING CURRENT MACHINES

Why alternating currents are used — the transformer — long-distance transmission — eddy currents — alternators — polyphase circuits — A. C. motors — rotating field — squirrel-cage rotor — A. C. power — wattmeters.

357. Why alternating currents are used. For heating and lighting an alternating current is just as satisfactory as a direct current. For plating and refining an alternating current cannot be used because a unidirectional current is necessary to make a metal deposit. If motors are to be run by an alternating current, a special type of motor is generally used, which is quite different from the ordinary direct-current motor. The real advantage in the use of alternating currents is **economy of transmission**. This is made possible by a simple and efficient machine known as a **transformer**.

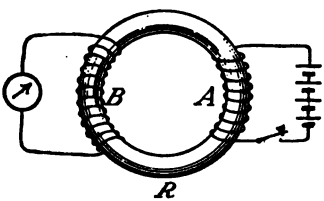


FIG. 319. — Faraday's ring transformer.

358. Induced currents in a transformer. As long ago as 1831 Faraday wound two coils of wire on a soft iron ring, as shown in figure 319. When coil A was connected with a battery and coil B with a galvanometer, he found that the needle of the galvanometer was disturbed every time the circuit was made and every time it was broken.

The modern transformer consists of two coils side by side on a common iron core not unlike Faraday's ring. When an

alternating current is set up in one coil, called the **primary**, it magnetizes the iron core, causing surges of magnetic flux, first in one direction and then in the opposite direction. Since this magnetic flux passes through the second coil, called the **secondary**, as well as the first, it induces an alternating current in the secondary. Since the same number of lines of force pass through both coils, *the volts per turn are the same. Therefore the total voltage in the primary coil is to the total voltage in the secondary coil as the number of turns in the primary is to the number of turns in the secondary.*

The line voltage in the street is often 2200 volts, which is too high to be safely used in private houses. It is therefore necessary to transform or "step down" to 110 volts. A primary coil of fine wire is connected to the 2200 volt circuit, and a secondary coil of coarse wire is connected with the lamp circuit of the house. The primary coil must have 20 times as many turns as the secondary. The secondary coil must be made of larger wire than the primary coil, because the secondary current is about twenty times the current taken by the primary. Thus the transformer delivers the same amount of energy which it receives, except for a small amount (from 2 % to 5 %), which is lost as heat in the transformer. The efficiency of a transformer is therefore very high, from 95 % to 98 %.

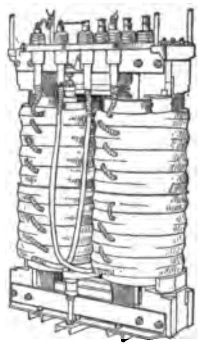


FIG. 320. — Core type of transformer.

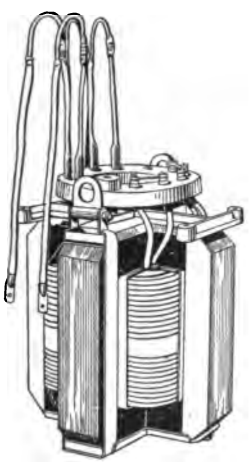


FIG. 321. — Shell type of transformer.

359. Commercial forms of transformer. Transformers are built in two general types: (*a*) the **core type** (Fig. 320), in which the coils are wound around two sides of a rectangular iron core, and (*b*) the **shell type** (Fig. 321), in

which the iron core is built around the coils. The iron core of both types is made of sheets of mild steel. To keep the coils insulated, the transformer is put in an iron case and surrounded with oil. These iron cases (Fig. 322) are commonly attached to poles near houses wherever the alternating current is used for lighting purposes.

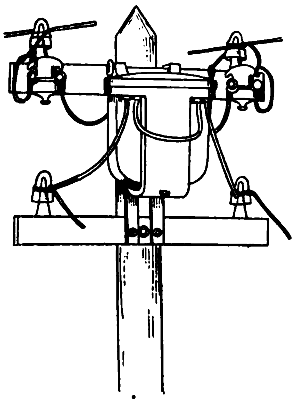


FIG. 322. — Transformer case mounted on pole.

360. Uses of transformers. In electric light stations it is common practice to use alternators to generate electricity at 2200 volts. The current is transmitted at this high voltage to the various districts, where it is transformed or "stepped down" to 110 volts for use in lighting houses. Another important use of the transformer is to furnish large currents at very low voltage for electric furnaces and electric welding.

To illustrate this, we may wind a turn or two of very large copper wire around the core of a small step-down transformer (Fig. 323), and connect its primary to a 110 volt A.C. circuit (if one is available). The ends of the large wire should be attached to a couple of iron nails. If, when the current is on, the tips of the nails are brought together, they get red hot and can be welded.

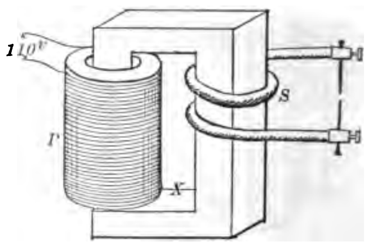


FIG. 323. — Step-down transformer, used for welding.

The adjoining rails of a car track are often welded together in this way. A heavy current is required for a short time, and is obtained by using a step-down transformer, in which the secondary consists of only one or two turns, made of very large copper bars. The ends

of this secondary are clamped to the rails to be welded, one on each side of the junction.

361. Long-distance transmission of power. By the use of alternating currents of high voltage, even up to 100,000 volts, power is now transmitted very long distances. For example, electric power is generated in hydroelectric power plants in the Sierra Nevada Mountains of California, and transmitted 200 miles to San Francisco. To understand why the economical transmission of electricity demands such high voltage, we have only to recall that the power transmitted is the product of the voltage and the current strength. Evidently, then, if we can make the voltage high, the current can be low. But a smaller current means smaller losses in transmission, for they are due to the heating effect of the electric current, and we have already seen that this varies as the *square* of the current.

It is an impressive sight to see three or six copper cables, each about $\frac{3}{4}$ of an inch in diameter, suspended about 75 feet above the ground on steel towers (Fig. 324, opposite page 347), and to know that those wires are carrying 30,000 H. P. of electrical energy. Hydroelectric power plants are being developed all over the country. For example, at Niagara power plants are generating electricity, raising the voltage to 60,000 and transmitting some of the enormous energy available at the Falls to distant cities like Buffalo, Rochester, and Syracuse. Just outside the city limits there are substations where the voltage is reduced to about 2000, and then it is distributed to factories and for general use in lighting and traction. Before the current actually enters the buildings, the voltage is again stepped down to 220 or 110 volts.

362. Eddy currents. We have seen that the cores of transformers are made of soft sheet-iron, or "mild steel," stamped out in the desired shape and then assembled. In the construction of induction coils the cores are made of

soft iron wires which are put together in a bundle. If we examine the armature of a dynamo, we find that the iron drum is made of laminæ (sheets) of mild steel which are stamped out in the shape of disks with notches around the edge (Fig. 325), and then assembled on a framework called the "spider," and mounted on the shaft. In all these cases the sheets or wires are insulated from each other by a coating of shellac, which eliminates what are sometimes called Foucault or eddy currents.

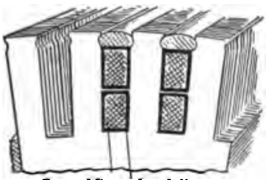


FIG. 325. — Laminated core of a dynamo armature.

We have already seen, in studying the generator, that when any conductor cuts lines of force, an induced electromotive force tends to send a current along the conductor. In the generator copper wires are provided to carry this current; but these wires are wound on an iron core, and if this core is itself an electrical conductor, an induced e. m. f. will be set up in it as it revolves in the magnetic field. This induced e. m. f. would send electric currents through certain portions of the core. These so-called eddy currents would soon heat the core, and would also retard the motion of the armature and waste power. To reduce these currents to as small a value as possible, the core is laminated in such a way that the insulation is transverse to the direction in which the eddy currents tend to flow.

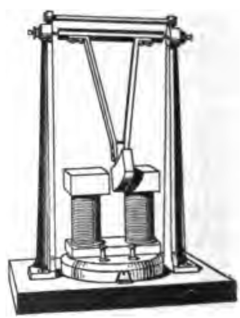


FIG. 326. — Damping by eddy currents.

363. Use of eddy currents in damping. To show that eddy currents tend to retard the motion of a conductor in a magnetic field, we may set up between the poles of a strong electromagnet a pendulum made of thick sheet copper (Fig. 326). If the magnet is not excited, the pendulum swings back and forth as any pendulum

does, but when we throw on the current in the magnet, the copper pendulum cannot swing through the magnetic field, and is instantly checked. The eddy currents set up in the copper tend to retard the motion of the pendulum, much as if it were swinging in thick sirup.

This effect is very useful in stopping the vibrations of the moving coil of a d'Arsonval galvanometer (section 274). The wire is usually wound on a light copper or aluminum frame, and the eddy currents in this metal frame check its swinging. Such a galvanometer is called "dead-beat." We shall see, in section 372, that this same principle is used to check the rotation of a wattmeter.

Questions and Problems

1. What limits the voltage which it is practicable to use on high-tension transmission lines?

2. Why are the cables for long-distance transmission sometimes made of aluminum instead of copper?

3. If a step-down transformer is to be used to change the voltage from 1100 to 220, what must be the ratio of turns of wire on the primary and secondary coils?

4. A transformer has 1000 turns on the primary and 50 turns on the secondary and the primary current is 20 amperes. About how much is the secondary current?

5. What generates the heat required to weld the nails in the experiment shown in figure 323? Why does not the copper wire *S* melt as well as the the tips of the nails?

364. Alternators. When a coil of wire is rotated in a magnetic field, we have seen (section 319) that the current changes its direction every half turn. That is, there are two **alternations** for each revolution in a bipolar machine. In a D. C. generator this alternating current is rectified by the use of a commutator. In the alternating current (A. C.) generator, called an **alternator**, the current induced in the armature is led out through slip rings, or collecting rings, as shown in figure 275. So almost any direct-current generator can be made into an alternator by substituting slip rings for the commutator.

The field magnet of an alternator is usually an electromagnet which is excited by direct current from a small auxiliary generator called the **exciter.**

Since it is only the *relative* motion of the armature windings and field magnet which is essential in any generator, large alternators are usually built with a **stationary armature** and a **revolving field.** The revolving projecting poles (*N*, *S*, *N*, *S*, in figure 327) sweep past the armature wires which are placed in slots around the inner periphery of the stationary structure *A*. The direct current for exciting the field coils is led in through brushes which rub on two insulated metal rings. The alternating current is led directly from the windings of the stationary armature through cables to the switchboard. Figures 328, 329, and 330 (opposite pages 364 and 365) show the revolving field and stationary armature of commercial machines of this type, and an assembled machine.

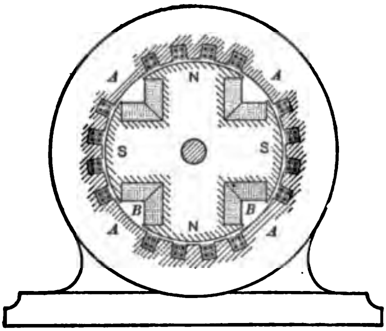


FIG. 327. — Revolving field and stationary armature.

365. Cycles and phase of alternating currents. When a conductor is moved past a magnetic *N*-pole, the induced e. m. f. is in one direction, and when it moves past an *S*-pole, the induced e. m. f. is in the opposite direction. This can be best represented by the curved line shown in figure 331. One complete wave is produced when

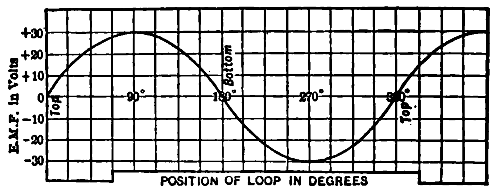


FIG. 331. — Alternating e. m. f. one complete cycle.

FIG. 328 *a*. — Revolving field for large slow-speed alternator. It has 36 poles.

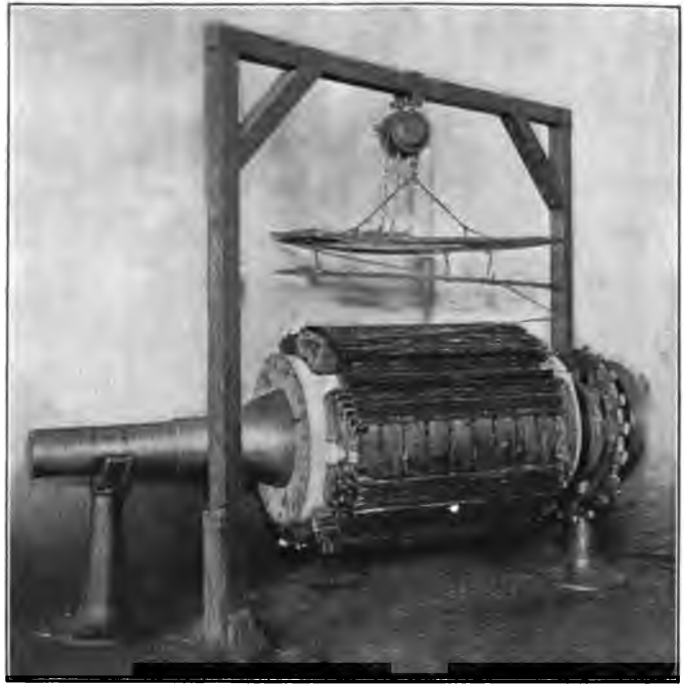
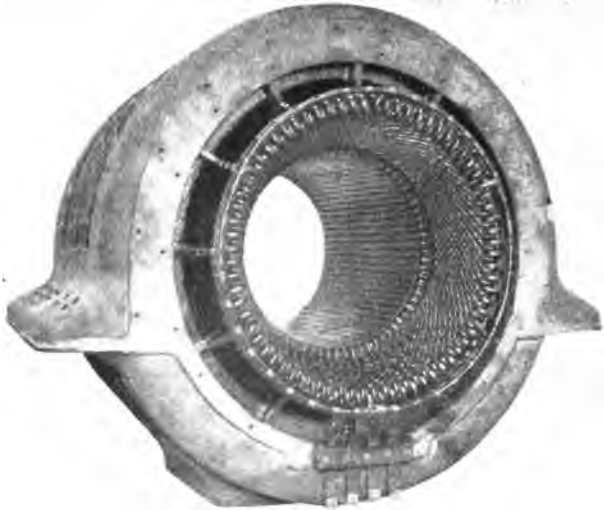


FIG. 328 *b*. — Partly wound revolving field for high-speed alternator to be driven by a steam turbine. It has 4 poles.



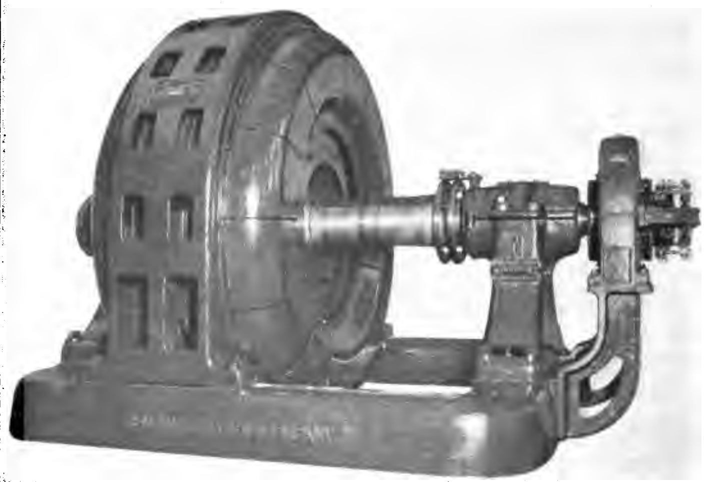


FIG. 329 (above). —Stationary armature of alternator.

FIG. 330 (below). — Alternator, belt driven. The long shaft allows the armature to be slid to one side so that the rotor can be examined and repaired. The small "exciter" is on the end of the shaft at the right.

a wire moves through a complete revolution in a bipolar machine, or from a north pole past a south pole to the next north pole in a multipolar machine, and is called a **cycle.**

In practice it is common to use for lighting an alternating current whose **frequency** is 60 cycles per second, while for power purposes 25 or even 15 cycle currents are common.

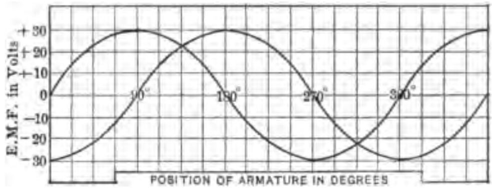


Fig. 332.—Two alternating currents which differ in phase.

A complete wave or cycle is called 360 **electrical degrees** by analogy with the complete revolution of a bipolar generator. Any point or position in the cycle is spoken of as a certain **phase.** When, for example, the cycle is half completed, the phase is said to be 180 degrees, and when the cycle is one fourth completed, the phase is 90 degrees. Two alternating currents of electricity, flowing in branch circuits, may be at different phases, as represented in figure 332, where one curve represents the current in one branch and the other curve the current in the other branch. In the case shown, one current is said to **lag** behind the other by 90 degrees.

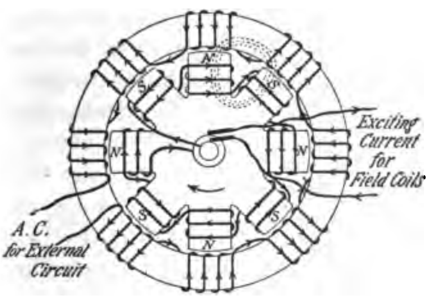


Fig. 333.—Diagram to represent armature and field coils on alternator.

366. Single and polyphase circuits. If we connect all the stationary armature coils of a generator in series, and revolve the field as shown in figure 333, a **single-phase** alternating current is produced whose *frequency* we can

determine by *multiplying the number of revolutions per second of the rotor by the number of pairs of poles.* To make use of this current for any purpose, such as electric lighting, we have simply to cut this armature circuit at any convenient

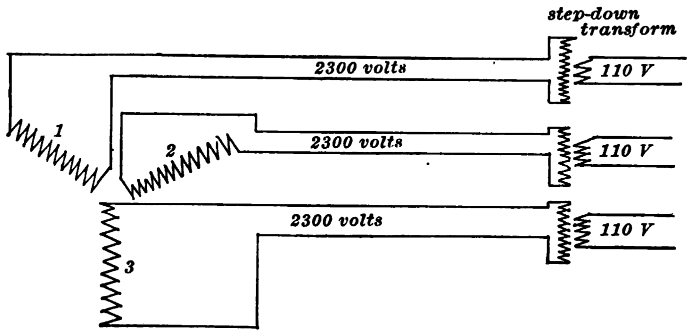


FIG. 334. — Alternating system three phases and six wires.

point and connect the ends directly to the mains. It will be noticed that there are as many coils on the armature as there are poles in the field magnet in the single-phase machine.

It has been found more economical of space to have more than one coil for each pole of the field, and so we have **two-phase** and **three-phase** machines, in which there are two or three sets of coils on the armature. In the three-phase machine, which is the type most used to-day, the three sets of armature coils may each be used separately to furnish electricity for three separate lighting circuits, as shown in figure 334.

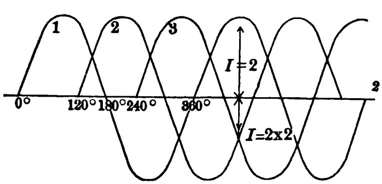


FIG. 335. — Curves for three-phase current system.

The currents in the three circuits differ in phase by 120 degrees (Fig. 335). It will be seen that the currents are such that at any instant their sum is zero.

To save wire, electrical engineers have devised ways of connecting the three sets of coils so as to have only three-line wires, instead of six, as shown in figure 336.

367. Use of alternators. The revolving-armature type of alternator is generally used only in small electric lighting stations. Large alternators of the revolving-field type are usually mounted on the same shaft (direct-connected) with the driving engine or water wheel. Alternators of very large capacity are now extensively used with steam turbines. They can be comparatively small in size because they are driven at such high speed. These alternators have a revolving field of only a few poles (sometimes only two) and a wide air gap between the armature core and the field poles. Figure 337 shows a 7500 kilowatt alternator mounted on the crank shaft of a 10,000 horse-power steam turbine, having a speed of 1800 revolutions per minute. In high-tension transmission, the three-wire three-phase system is commonly used.

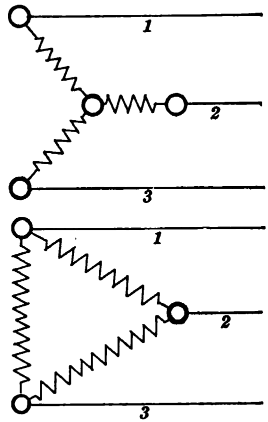


FIG. 336. — Y and Δ connections on three-phase circuit.

QUESTIONS

1. How can the engineer at the power house control the frequency of an alternating current?
2. How many revolutions per minute will an 8-pole machine have to make to give a 60-cycle current?
3. What objection is there to using a 15-cycle current for lighting purposes?
4. Draw a diagram to show two alternating currents which differ in phase by 45 degrees.
5. How much do the two currents generated by a two-phase alternator differ in phase?

368. Alternating current motors. An A. C. generator can be run as a motor, provided it is first brought up to the exact speed of the alternator which is supplying current to it and put in step with the alternations of the current supplied. Such a machine is called a **synchronous motor.** Since it is not self-starting, it is not convenient for general use, but is used in substations to drive D. C. generators.

An ordinary series motor, by certain modifications in its design, can be made to operate on either D. C. or A. C. systems. These so-called A. C. **commutator motors** or **single-phase series motors** are coming into use for electric cars and locomotives when an alternating current is used. They are also to be found, in very small sizes, on egg-shaking machines in drug stores, and on vacuum cleaners. They are labeled "A.C. or D.C." on the name plates.

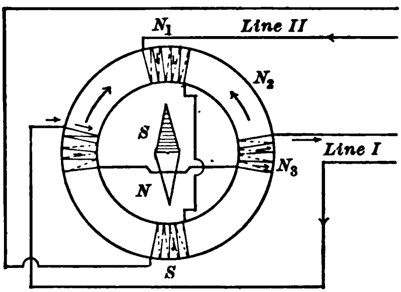


FIG. 338. — Iron ring excited by two currents 90 degrees apart.

The A. C. motor most frequently used is the **induction motor.** The distinctive features of this motor are that the stationary winding, or "stator," sets up a rotating magnetic field, and that the rotating part of the motor, or "rotor," is built on the plan of a squirrel cage. These will be discussed in turn.

369. Rotating magnetic field. To produce a rotating field, we will suppose that we have two alternating currents of the same frequency, but differing in phase by 90 degrees, and that we connect them to two sets of coils wound on a ring, as shown in figure 338.

When the current in line I is at a maximum, it will be seen from the curves (Fig. 339) that the current in line II is zero. The top of the ring is therefore a north pole, N_1, and

the bottom is a south pole, S. One eighth of a cycle (45 degrees) later, current 1 has decreased in strength and current 2 has increased in strength. The result of both currents is to form a north pole in the position N_2, 45 degrees farther along. One eighth of a cycle (45 degrees) later, current I has dropped to zero and current II is at a maximum. This brings the north pole of the ring to the right side (N_3). Evidently the north pole is traveling around the ring, and will make a complete circuit for each complete cycle of the current. This produces a **rotating field**, and would cause a magnet, such as NS, to rotate with the field. We should then have a little two-phase A. C. motor.

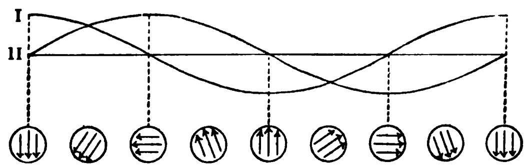


Fig. 339.—Curves of two alternating currents, which differ in phase by 90 degrees.

Figure 340 shows a working model to demonstrate the rotating field produced by a two-phase current system.

370. The rotor of an induction motor. The rotating magnet can, of course, be replaced by an electromagnet, which

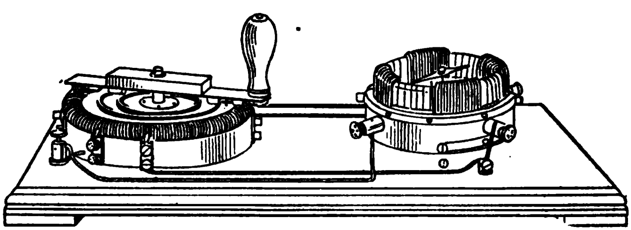


Fig. 340.—Working model of two-phase rotating field.

is excited by some outside source of direct current. The rotor of a commercial A.C. motor is, however, much simpler, It consists of an iron core, much like the core of a drum

armature, with large copper bars placed in slots around the circumference and connected at both ends to heavy copper rings. This is called a **squirrel-cage rotor** (Fig. 342).

When it is placed in a rotating magnetic field, the conductors on the two sides and the rings across the ends act like a closed loop of wire, and a large current is induced, even though the rotor has no electrical connection with any outside circuit. This large induced current makes a magnet of the iron core, and the field, acting on this magnet, drags it around.

The rotor can never spin quite as fast as the magnetic field. If it did, there would be no cutting of lines of force, no currents would be induced, and there would be no power available to drive the rotor against its load.

As in the case of the Gramme ring dynamo, the ring winding is not used in practical motors. The common construction is to slip coils into slots in the inner periphery of a laminated iron "stator," as shown in figure 341. A squirrel-cage rotor (Fig. 342) is simple and strong, and needs only to be kept cool. This is done by air circulated through the core by fan blades. The assembled machine (Fig. 343) is simple, strong, compact, and almost "fool-proof." For these reasons, the induction motor is finding a wide field of usefulness in shops and factories, and even on electric locomotives.

371. A. C. power. We have seen that we can determine the power of a direct-current circuit by multiplying the volts and amperes together. With a non-inductive circuit, such as a lamp, we can do the same with alternating currents. In the case of machines which have self-induction, that is, coils of wire with iron cores, the number of **volt amperes** is **greater** than the true number of **watts**. Although we cannot attempt to show in this book just how the watts may be computed from the volts and amperes of an alternating current, yet we can see why it is not a simple case of multiplication.

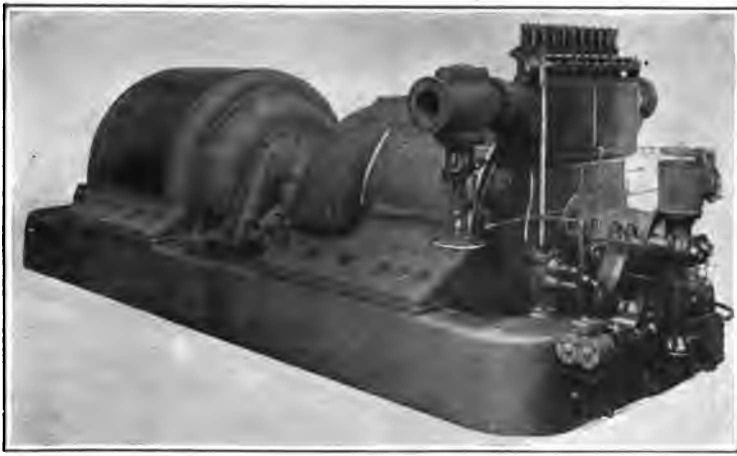


Fig. 337. — 7500 k.w. alternator driven by steam turbine.

Fig. 341. — Stator of an induction motor.

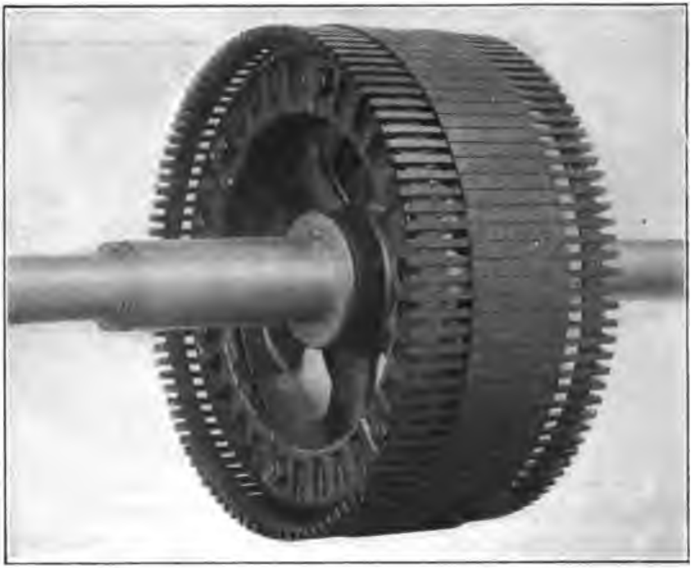


Fig. 342. — Squirrel-cage rotor of induction motor.

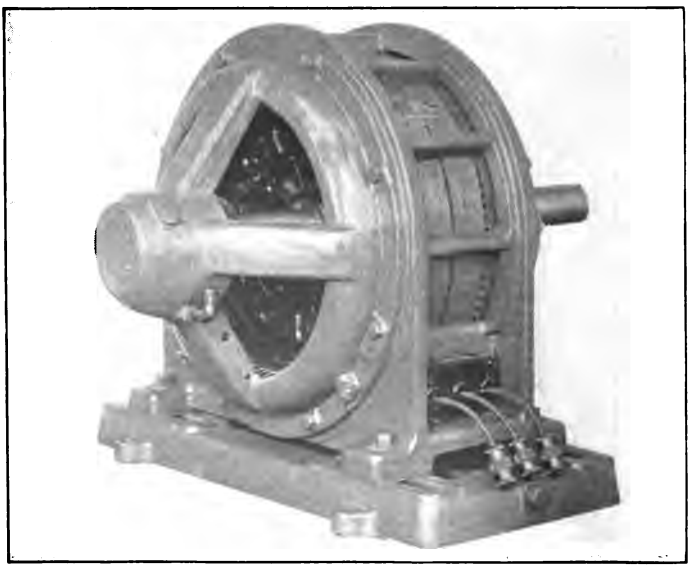


Fig. 343. — Induction motor.

In the first place, we will represent the varying electromotive force by the "pressure" curve in figure 344, and the varying current by the current curve in the same figure. It will be noticed that the current curve *lags*, that is, it starts after the e. m. f. curve. This is due to the self-induction of the circuit which impedes the flow of the alternating current. To get the power of such a current, we should have to multiply the *simultaneous* values of current and pressure for a great number of points, and then get a general average of these products.

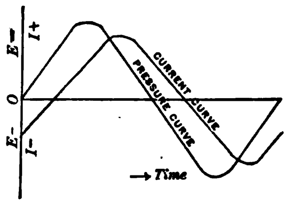


FIG. 344. — Voltage and current curves.

Now what an A.C. ammeter (which we will not attempt to describe) records, is the "effective value" of the alternating current; that is, the value in amperes of the direct current which would produce the same heating effect. It can be shown that this "effective value" of an alternating current is about 0.7 of its maximum value. The effective value of an electromotive force is said to be one volt, when it will develop an alternating current of one ampere in a non-inductive resistance of one ohm. It is also about 0.7 of the maximum value of the e. m. f. Evidently it will not do to multiply these effective values of current and voltage together, because, in the averaging process described above, large values of the current are likely to be paired with small values of the e. m. f., and *vice versa.*

It can be shown that the **A. C. watts** are equal to the **volt-amperes** times a factor, which is called the **power factor.** This factor varies according to the circuit. It is always less than 1 for an inductive circuit.

372. Wattmeters. Every user of electricity should be interested in the **recording wattmeter**, which records on dials, like those of a gas meter, the number of kilowatt hours of electricity consumed. It is on the readings of this instrument

that the monthly bills are based. Figure 345 shows the Thomson form of wattmeter. It is really a little shunt motor, the armature of which turns at a speed proportional to the rate at which electrical energy is passing through it. This armature is geared to the recording dials. The field of the instrument is made by stationary field coils which are connected in series with the line. The field strength is

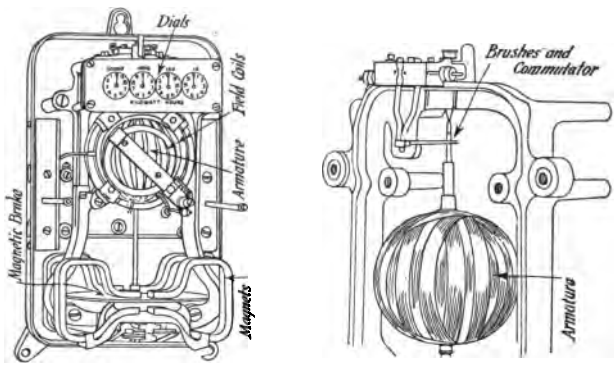


FIG. 345. — Thomson's watt-hour meter.

therefore proportional to the current flowing in the main line. The armature is connected across the line, and takes a current proportional to the voltage across the line. Therefore, the torque which turns the armature is proportional to the product of the current and the voltage; that is, to the watts in the line.

The inertia of such a machine would make it run too fast, or fail to stop when the current stopped, if it were not for the electric damping caused by the rotation of an aluminum disk between the poles of permanent magnets. The eddy currents generated in the disk tend to retard its motion.

This type of wattmeter is used for both A. C. and D. C. work. When used with alternating currents it automatically averages the products mentioned at the top of page 371.

SUMMARY OF PRINCIPLES IN CHAPTER XIX

In a transformer:—

$$\frac{\text{Voltage on primary}}{\text{Voltage on secondary}} = \frac{\text{turns of primary}}{\text{turns of secondary}}.$$

In an alternator:—

Frequency = revolutions per second × number of pairs of poles.

A. C. power = amperes × volts × power factor.
Power factor usually less than one.

QUESTIONS

1. The iron case of a transformer is often corrugated. Why?

2. Why must the dielectric strength of the oil used in transformers be carefully tested?

3. In long-distance transmission of power by high-tension lines, the wires are often supported on steel towers 50 feet or more above the ground, and the company gets a right of way to a strip of land 100 feet wide over which to run its wires. Why these precautions?

4. What is gained by making the armature of big alternators stationary, and rotating the field?

CHAPTER XX

SOUND

What makes sound — what carries sound — velocity of sound — water waves — velocity, wave length, and frequency — longitudinal waves — sound waves — loudness and distance — directing sound — reflecting sound — musical tones — intensity, pitch, and quality — resonators — overtones — beats — the musical scale — stringed instruments — wind instruments — membranes — the phonograph.

373. What makes sound? When a bell rings, we see the hammer or clapper hit the bell, and hear the sound which it makes. If we hold a pencil against the edge of the bell just after it has been struck, we find that the metal is moving to and fro very rapidly. When a guitar string is plucked, it gives forth a note which we can hear, and at the same time we can see that the string looks broader than when at rest. We conclude that the string is vibrating or oscillating back and forth. When we strike a tuning fork and hold it near the ear, we hear a note, and if we touch the fork to the lips, we feel its vibratory motion.

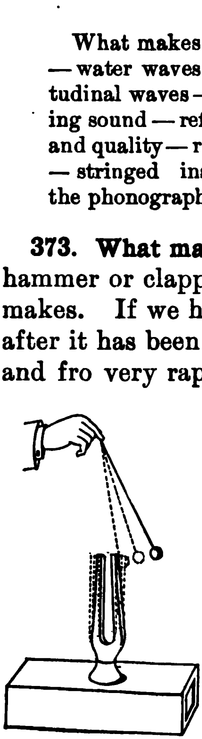


FIG. 346. — Vibration of tuning fork made visible.

To make visible the vibration of a tuning fork let us touch it to a light glass bubble suspended on a thread (Fig. 346). The bubble is set violently in motion.

Another way to show the vibratory motion of a fork is to attach a point of stiff paper to one prong. Let us set such a fork in vibration and draw it over a piece of smoked glass (Fig. 347). The curve which is traced is easily made visible by putting white paper behind the glass.

Whenever we look for the source of a sound, we find that something has been set in motion. It may be that something has fallen, a bell has been struck, a whistle has been blown, or some one has shouted; always something has been set vibrating which has caused the sensation of sound.

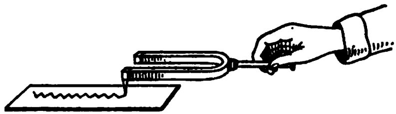


FIG. 347. — Curve traced by a vibrating fork.

374. What carries sound? Ordinarily the air, which is everywhere about us, brings sound to our ears. To make this evident let us try the following experiment.

Let us suspend an electric bell under the receiver of a good vacuum pump, as shown in figure 348. If we set the bell to ringing and then pump out the air, we find that the sounds become fainter and fainter. When we let the air in again, the bell sounds as loud as at first. It seems probable that the bell would become quite inaudible if we could get a perfect vacuum, and if no sound were conducted out by the suspension wires.

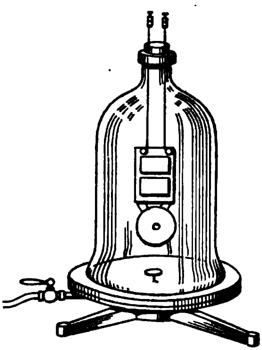


FIG. 348. — Sound is not carried through a vacuum.

We know that both heat and light can traverse a vacuum, as in the case of the electric incandescent light bulb, but we see from the last experiment that sound does not traverse a vacuum.

It can be shown that other gases besides air carry sound, and that liquids and solids are even better carriers of sound than gases. For example, if one holds his ear under water while some one hits two stones together at some distance away, the sound is heard very distinctly. It is also a familiar fact that one can hear a train a long distance away by putting one's ear close to the steel rail. Loud sounds, like those of cannon, or of volcanic eruptions, can be heard at a

distance of several hundred miles by putting one's ear to the ground.

To show that liquids transmit sound, let us put the stem of a tuning fork into a hole bored in a large cork. If we set the fork in vibration, it is hardly audible; but if we hold it with the cork resting on the surface of a glass of water, we hear it distinctly. The sounds seem to be coming from the table on which the tumbler of water stands. This experiment shows that the vibration of the tuning fork is transmitted through the cork and the water to the air in the room.

To show that solids transmit sound, we may hold one end of a long wooden stick against a door, and rest a vibrating tuning fork on the other end; the sound of the fork seems to be coming from the door. The wooden stick here serves as the sound carrier and transmits the vibration of the fork to the door.

So we conclude that solids, liquids, and gases may serve as carriers of sound.

375. How fast does sound travel? In an ordinary room one is not aware that it takes any appreciable time for sound to travel from its source to one's ears; but in a large hall, or out doors, one often hears an echo, which shows that sound does take time to travel to a reflecting surface and back. During a thunder shower we hear the roll of the thunder after we see the flash. The farther away the lightning discharge is, the longer the interval between seeing the flash and hearing the rumble. Every one has doubtless *seen* the steam from a distant whistle, and then later *heard* the whistle. So there is no doubt that sound travels much more slowly than light.

One way to measure how fast sound travels is to discharge a cannon on a distant hill and measure the time between seeing the flash of the cannon and hearing its report. In one such experiment, which was performed by two Dutch scientists in 1823, the cannons were set up on two hills about eleven miles apart, and observations were made first from one hill and then from the other, to eliminate the error due to wind. They concluded that sound travels 1093 feet (or 333 meters) per second, which was remarkably near the truth, considering

the instruments they had. Since then, several men have made determinations of the velocity of sound in air, which show that at 0° C and 76 centimeters pressure the ***velocity of sound is 1087 feet (or 331 meters) per second.*** The speed of sound in water is about 4.5 times the speed in air, and in steel it is more than 15 times as great as in air. It has also been found that the speed of sound in air increases about 2 feet (or 0.6 meters) per second for each degree centigrade rise in temperature. For practical purposes it is enough to remember that ***sound travels about 1100 feet per second.***

Problems

(Assume that the time taken by light to travel ordinary distances is negligibly small)

1. The sound of a steam whistle is heard 2.6 seconds after the steam is seen. About how far away is the whistle?
2. A man can see the hammer strike a bell once every 2 seconds. If the man is a mile away, what is the interval between the sounds of each stroke?
3. On a hot summer day, when the temperature is 30° C, the flash of a gun is seen 2 miles away. How long after the flash will the report of the gun be heard?
4. A stone is dropped from the top of the Woolworth Building in New York, which is 750 feet high. How long before a man on top would hear the sound of the stone as it struck the pavement? (The time includes the time for the stone to fall and for the sound to return.)
5. If an experiment shows that sound travels in water 4814 feet per second at 14° C, how many times as fast does sound travel in water as in air at this temperature?

376. Sensation of sound. We have been considering the transmission of "sound" through gases, liquids, and solids, although we know that it is merely a sort of motion which is transmitted. Ordinarily we find it hard to think of sound without thinking of an ear to hear it. Thus we find people asking whether a waterfall in a very remote part of the earth, never visited by any man or animal, makes any

sound. Evidently there are two things which are called "sound" the **vibrations**, and the **sensation** they produce when they strike against the tympanum or eardrum. The study of what happens in the ear and brain is properly left to physiology and psychology. In physics we shall study only the vibrations in the air or other transmitting medium, and shall refer to them when we say "sound." In this sense the waterfall makes just as much sound whether there is an ear to hear it or not.

377. Sound a wave motion. Evidently nothing material (that is, weighable) travels from the source of a sound to the ear; otherwise, how did the sound of the electric bell under the bell jar get through the glass? This and other facts point unmistakably to the conclusion that what is transmitted is merely a vibration or **mode of motion**, called a **wave**.

378. Water waves. Since sound waves are usually invisible, we will start with a study of **water waves**. When a stone is dropped into a smooth pond, a disturbance is produced which extends over the surface of the water in circles centered at the place where the stone struck. The water is pushed down and aside by the stone, forming a circular ridge which expands into a larger circle, and is followed by a second circular ridge which expands, and so on. The result is that the surface is soon covered with a series of circular swells which are separated by circular troughs, all moving away from the center of the disturbance.

To study these water waves more carefully, let us pour water into a long tank with glass sides (Fig. 349) to a depth of 2 inches, set a paddle upright about 6 inches from one end of the tank, and start a wave by drawing the paddle to the end of the tank. It will be observed that the wave travels to the other end of the tank. There it is turned back or reflected, returning to the first end, undergoing another reflection, and so on. By measuring the length of the tank and observing the time of six round trips of a wave (observe the rise and fall of the water at one side) we can calculate the speed of the wave.

If we pour more water into the tank until the depth is 3 inches, and

again time six round trips and calculate the speed of the wave motion, we shall find that waves travel faster in deep water.

To study stationary water waves we place a little block on the water at one end of the tank. By raising and lowering the block periodically,

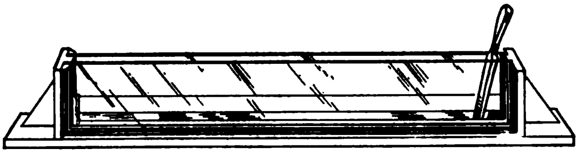


FIG. 349. — Tank for water waves.

we may set up stationary water waves, in which the water simply "seesaws" up and down with no apparent backward and forward motion.

The surface of a water wave may be represented by the curved line shown in figure 350. The stationary points, *A*, *B*, *C*, *D*, etc., are called the **nodes**; the intervening spaces are called the **loops**, or **internodes**. The water between nodes oscillates up and down; when it is up, it forms a crest, and when it is down, it is a trough. A crest and trough together form a wave, as from *A* to *C*, or *B* to *D*. The length of a wave (l) is measured horizontally from any point on one wave to the corresponding point in the next wave. Corresponding points are called points in the same **phase**. The amplitude (d) of the wave vibration is half the vertical distance from trough to crest.

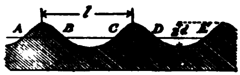


FIG. 350. — Surface of a water wave.

379. Relation between velocity, wave length, and frequency. In the case of the waves started by throwing a stone into a quiet pool, we know that while the circular waves grow larger and larger, any particular crest seems to move out radially until it reaches the bank or dies away. The distance which a crest travels in one second is called its **velocity**. The number of crests passing a fixed point in one second is called the **frequency**. The time it takes one wave to pass a

given point, that is, the time between crests, is called the **period** of the wave motion.

If n is the number of waves passing a given point in one second, that is, the *frequency*, and if p is the time required for one wave to pass a given point, that is, the *period*, then,

$$p = \frac{1}{n}.$$

Again, if l is the length of one wave in feet, and n is the number of waves passing any point in one second, the distance traveled by a wave in one second, that is, its velocity v in feet per second, is equal to n times l; that is,

$$v = nl.$$

It should be remembered that it is only the wave form that travels over the surface of the water, not the water particles themselves. Thus if we float a cork or a toy boat on a pool over whose surface waves are passing, the cork or boat merely bobs up and down as a wave passes, but is not carried along with it.

380. Transverse and longitudinal waves. An easy way of illustrating wave motion is to fasten one end of a piece of rubber tubing about 20 feet long to a hook in the wall. If we take the free end in the hand, we can, by a quick shake, send a wave along the tube (Fig. 351). If a single depression is sent along the tube to the fixed end, it is reflected and returns as an elevation; in like manner a single elevation sent along the tube comes back as a depression.

FIG. 351. — Waves in a rubber tube.

In the case of water waves and of the waves in a tube or cord, the particles of water or tubing oscillate up and down, while the disturbance moves horizontally. Such waves are called **transverse** waves.

A second kind of wave motion takes place in substances such as gases and wire springs, which are elastic and com-

pressible. This kind of wave can be studied by letting a coil of wire represent the substance through which such waves are transmitted.

Figure 352 represents a spring whose turns are large and are supported by threads. If we strike the spring at one end, we compress a few turns near that end. These move slightly and compress those just ahead, and

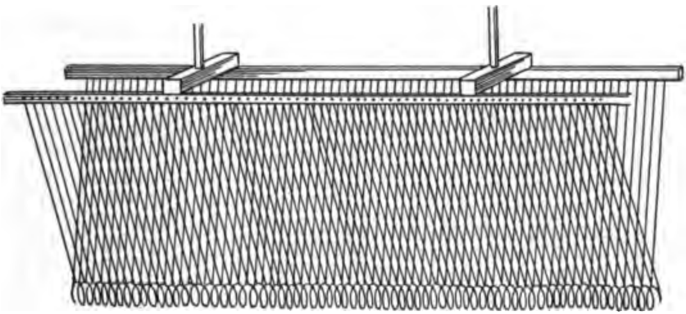


FIG. 352. — Spring wave model.

these in turn squeeze together the turns still farther along. Thus a pulse or wave goes along the spring.

Next let one end of the spring be given a quick pull, so that the turns near by are drawn apart for an instant. Then the adjacent turns will be pulled over, one after another, until this disturbance reaches the other end. Thus it is seen that any push or pull given to the spring at one end is transmitted as a push or pull to the other end.

Waves of this sort, in which the particles of the transmitting material move back and forth in the direction of the advance of the wave, are called compression or **longitudinal** waves.

381. Longitudinal vibration in solids. Not only springs, but gases and even solids like steel, transmit vibrations longitudinally.

If we clamp a steel rod in the middle and rub it lengthwise with a cloth dusted with rosin, a clear, ringing sound may be produced. That the rod has been set in vibration longitudinally can be shown by a little

ivory ball hung by a cord so as to rest against the end of the rod. When the rod is vibrating, the ball will swing violently out, as shown in figure 353.

Another mechanical illustration of the method by which a push or pull may travel a long distance, although the individual particles move only very minute distances, is shown in the following experiment: —

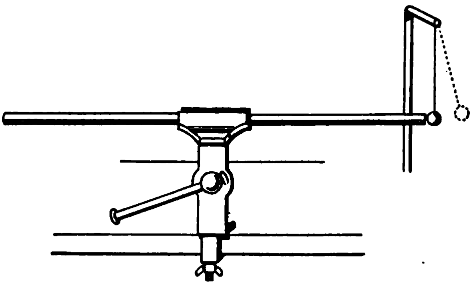


FIG. 353. — Ball driven from end of rod.

The apparatus shown in figure 354 consists of several glass-hard steel balls hung up in a line so that they just touch each other. If we pull aside the first ball and let it fly back and strike the line of balls, the ball it strikes does not seem to move, nor the next one. In fact none seem to be affected by the blow except the ball on the opposite end, which flies out about as far as the first ball fell.

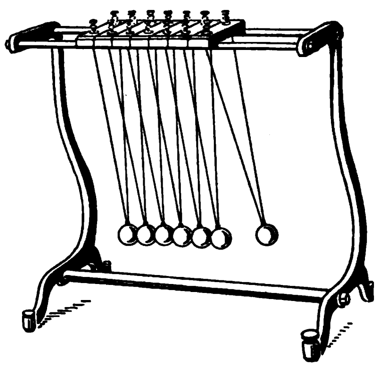


FIG. 354. — Illustrating how sound travels from particle to particle.

Since steel is very elastic, the impact of the first ball is handed along from ball to ball until it reaches the end one. It is as though a push were given to the first of a column of boys standing in line. It is transmitted along the line, and the last boy is pushed over.

382. Sound waves. We think of the air in sound waves as vibrating to and fro in the direction of propagation like the turns of the spring; that is, *sound waves are longitudinal or compression waves, made up of alternate condensations*

and rarefactions. Just as a stone thrown into a pool makes waves which spread out in ever widening concentric circles, so we think of a bell as sending out **spherical** waves. These are made up of alternate spherical shells of compressed and rarefied air, traveling out in every direction through space.

To form a picture of a sound wave traveling through a speaking tube, let us imagine that the spiral spring of the model (Fig. 352) is replaced by a column of air, which has a

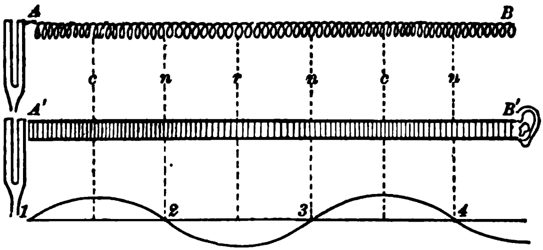


Fig. 355. — Diagram to show sound waves by a curve.

tuning fork at one end, giving little pulses to the air column, while an eardrum at the other end receives these pulses (Fig. 355).

The successive condensations and rarefactions of the air are indicated by *c* and *r* in *A*. The disturbance travels from the fork to the air, but the intervening air at any point merely oscillates a very little to and fro. The curve in figure 355 is a graphical representation of these sound waves, in which the crests, 1–2, 3–4, etc., represent condensations or compressions, and the troughs, 2–3, etc., represent rarefactions. The amplitude of the wave corresponds to the distance each particle of air moves to and fro from its original position. A sound wave includes a complete crest and trough, that is, a condensation and rarefaction, and the distance between two corresponding points in any two adjacent waves is called the wave length.

Since the same relation between velocity, wave length, and frequency holds for sound waves as for water waves, we can easily compute the length of a sound wave.

Suppose a tuning fork is giving 256 vibrations each second, and that the velocity of sound is 1120 feet per second. Then the length of each wave is 1120 feet divided by 256, or about 4.4 feet.

Or substituting in the wave equation,

$$v = nl,$$
$$1120 = 256\, l,$$
$$l = 4.4 \text{ feet.}$$

To picture a sound wave spreading through the open air, we may imagine a great number of spiral springs radiating out from a common center at the source of the sound, all receiving an impulse at the same time.

Problems

1. An *A* tuning fork on the "international scale" makes 435 vibrations per second. What is the length of the sound wave given out?
2. A vibrating string gives out sound waves 2 feet long. What is the frequency of the waves?
3. The period of a sound wave is found to be 0.0025 seconds. What is the length of the wave?
4. A bell whose frequency is 150 vibrations per second is sounded under water, in which sound travels at the rate of 4800 feet per second. Find the wave length produced by the bell.
5. If the highest tone which the ear can recognize makes 30,000 vibrations per second, what is the shortest wave which the ear appreciates?

383. Intensity or loudness of sound. It must always be remembered that when a bell is struck, the sound is heard in all directions, which means that sound waves spread out in all directions as shown in figure 356. As the distance from the source increases, the spherical waves spread out over more surface, and so the intensity of the sound decreases. For example, a bell 10 feet away will sound one fourth as loud -- the same bell 5 feet away, and if 15 feet away, it sounds

one ninth as loud as when 5 feet away. This is because the energy of the wave must be imparted to nine times as many particles at a distance of 15 feet as at a distance of 5 feet. *In general, the intensity of sound varies inversely as the square of the distance.*

If one ascends to a high altitude, as on a mountain top or in a balloon or aëroplane, the air becomes less dense and so not so good a carrier of sound. This makes it difficult to transmit sounds. *In general, the intensity of sound depends on the density of the medium through which the sound is transmitted.*

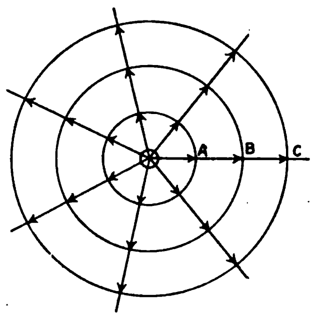


FIG. 356. — Sound waves spread out in all directions from this source.

384. Speaking tubes and megaphones. The **speaking tubes** used to connect rooms in buildings and ships serve to prevent the spreading out of sound waves in all directions, and so the sound is heard with almost its original intensity at the distant point. Sharp bends in such tubes should be avoided, as they cause reflected waves, which run back.

In the **megaphone** the sound waves which come from the mouth are not permitted by the walls of the instrument to spread out in all directions. In this way the energy of the voice is sent largely in one direction.

385. Reflection of sound. Just as any elastic body like a rubber ball bounds back when thrown against a brick wall, or a water wave is turned back by a stone enbankment, so a sound wave is turned back or **reflected** when it strikes against another body, such as a building, cliff, or wooded hillside, or even a cloud. The returning wave is called an **echo**. If the reflecting wall is near, as in a closed room, one may hear an echo almost at the same instant as the sound. This confuses

a hearer, and is an acoustical defect in the room. It can often be remedied by putting an absorbing material on the reflecting wall. When the reflecting surface is 25 or more yards distant, the echo is distinct from the original sound, and excites interest and curiosity. The greater the distance, the longer is the time before the reflected wave strikes the ear, and therefore the more distinct the echo becomes. When we have parallel walls, as in a narrow cañon, or objects at different distances, the echo is multiple or repeated, which means that the same sound is heard several times. For example, the roll of thunder results in part from the reflection of the sound from a succession of mountains or clouds.

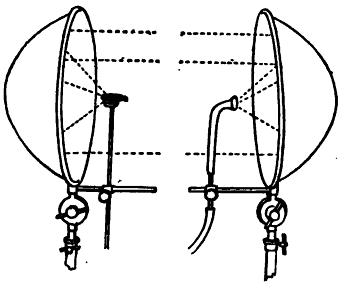


FIG. 357.— Sound of a watch reflected by mirrors.

The following experiment shows that sound waves, like light waves, are reflected by curved surfaces. If two large parabolic mirrors face each other, as in figure 357, a watch at the principal focus of one mirror can be distinctly heard across the room by holding an ear trumpet at the focus of the other mirror.

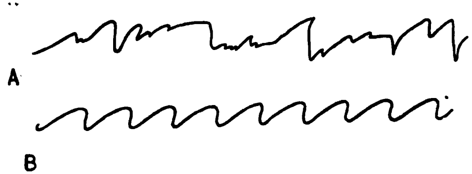


FIG. 358.—Curves to represent (*A*) noise and (*B*) music.

In buildings with arched ceilings it is sometimes possible to hear a whisper at a very distant place in the room because the sound is reflected from the ceiling and concentrated at the ear of the listener.

386. Musical sounds and noises. We all recognize some sounds, such as the slamming of a door or the rumbling of a wagon over cobblestones, as **noises**; while we recognize the

sounds from a piano wire or an organ pipe as **musical sounds** or **tones**. The difference between these kinds of sounds can be best expressed by the curves in figure 358, where *A* is the curve of a noise, and *B* the curve of a musical note.

It will be seen from these curves that a noise makes a very irregular and haphazard curve, while a musical note makes a uniform and regular curve. The latter produces an agreeable sensation on the ear, while the former makes a disagreeable sensation. The great German scientist, Helmholtz, expressed this distinction by saying, "The sensation of a musical tone is due to a rapid periodic motion of a sounding body; the sensation of a noise to a non-periodic motion."

387. Three characteristics of a musical note. A musical sound or tone has intensity or loudness, pitch, and quality or timbre, and each of these characteristics depends upon some physical property of the sound wave. The **intensity** of a sound depends on the amplitude of the vibration; the **pitch** depends on the frequency of the waves; and the **quality** depends on the vibration form.

388. Intensity. We have already seen that the intensity of sound in general diminishes as the distance of the ear from the source of the sound increases and also as the density of the air diminishes. The **intensity** of a musical sound for a given ear and at a given distance depends on the **amplitude of vibration** of the waves sent out. For example, a piano string or a tuning fork gives a louder sound when struck hard than when struck gently.

389. Pitch. When we speak of a musical note as high or low, we refer to its pitch. When we strike the keys of a piano in succession, beginning at one end of the keyboard, we recognize the difference in the tones produced as a difference in pitch. By holding a card against the teeth of a rapidly revolving wheel (Fig. 359) we can show that the **pitch** of the note produced depends on the **number of vibrations** per second; that is, upon the frequency of the vibrations.

We can show this very clearly by means of a **siren**. This is a metal disk (Fig. 360) with holes equally spaced around the edge, which can be rotated by some sort of whirling apparatus. If a current of air is directed through a tube against the holes, the regular succession of puffs produces a musical tone. As we increase the velocity of the wheel, the tone becomes higher; that is, its pitch is raised.

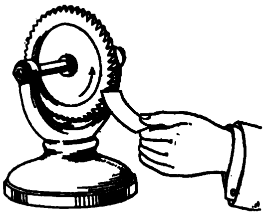


FIG. 359. — The pitch varies with the speed.

One way to measure the frequency of vibration of a musical tone is by means of such a rotating disk. Suppose the disk has 80 holes, and is attached to a motor making 1800 revolutions per minute. Since the disk makes 30 revolutions per second, there are $30 \times 80 = 2400$ puffs per second. The frequency of the tone emitted would be 2400 vibrations per second. This would be a rather shrill note. A standard *A* tuning fork makes only 435 vibrations per second.

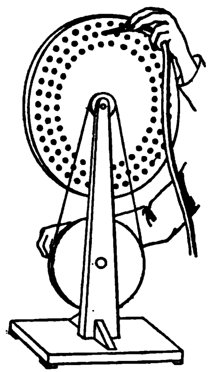


FIG. 360. — Pitch depends upon the rate of vibration.

390. Limits of audibility. The lowest tone which the human ear can recognize as a musical tone has a frequency of about 16 vibrations per second. If the sound has a frequency above a certain number, the ear does not recognize it at all. This upper limit of audibility varies with different people from 20,000 to 40,000 vibrations per second. A young person can usually recognize sounds of a higher pitch than an older person. In fact this is one of the evidences of the impairment of hearing with advancing age.

391. Quality or timbre. The third characteristic of a musical note is its **quality**. It is quality which enables us to distinguish between notes of the same pitch and intensity as produced by different instruments or sung by different voices. Even the same kind of instrument may produce

11. There is an old saying that "if you can count three between a flash of lightning and its thunder clap, the storm is not dangerously near." According to this how far away must the thunder cloud be for safety?

12. Explain just why the resonance experiment described in section 393 will not work if the length of the air column is half a wave length.

13. Explain how sound is produced by some form of automobile horn or signal in common use.

notes of different quality. For example, it is the quality of the tones produced by two violins which makes the great difference in their value. We recognize the voice of a friend over the telephone by its quality.

Helmholtz (1821–1894) first discovered the cause of these subtle differences in musical tones, which are called quality. In this investigation he made use of **resonators** which vibrated in sympathy with the tones to be studied.

392. Sympathetic vibrations. Every one has learned by experience how easy it is to set a swing vibrating by a succession of gentle pushes applied at just the right time, so that each push helps rather than hinders the swinging. Mere random pushes, on the other hand, accomplish very little. In much the same way sound waves or other slight impulses may set up strong vibrations in a body if they are timed to correspond exactly to its natural frequency of vibration. This is called **sympathetic vibration.**

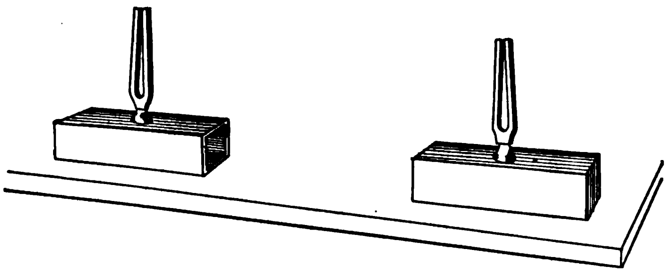


FIG. 361.—Sympathetic vibration of forks of the same pitch.

It can be strikingly shown by holding down the loud pedal of a piano, so that the dampers are lifted from the strings, and singing a clear, strong tone into the instrument. After the voice is silent, the sound is returned by the strings with enough fidelity to make the effect almost startling.

Another way to illustrate sympathetic vibrations is to put two tuning forks of the same pitch several feet apart (Fig. 361). If we strike one fork vigorously with a soft mallet, and then quickly stop it with the hand, the other will be heard even in a large room. It has been set in motion

by the sound waves from the first fork. If we change the pitch of one fork by sticking a bit of beeswax on one prong, the forks will be thrown slightly out of unison and will no longer respond to each other.

From this experiment it is evident that two tuning forks must vibrate at exactly the same rate to vibrate in sympathy. Certain articles of furniture and of glassware have definite rates of vibration of their own, and are set vibrating sympathetically when their particular note is sounded. It is the cumulative effect of feeble impulses repeated many times at regular intervals which sets up this sympathetic vibration.

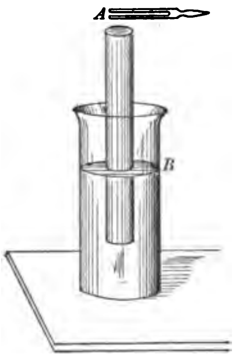


FIG. 362. — Reënforcement of a sound by an air column.

393. Resonators. That property of a sounding body which enables it to take up the vibrations of another body by sympathy, and to vibrate in unison with it, is called **resonance**. In the last experiment each tuning fork stood on a wood box open at one end and so constructed that the air column within the box has the same rate of vibration as the fork itself. Such an air column is called a **resonator**. It was the resonator rather than the fork itself that picked up the vibrations.

To show resonance, we may raise and lower the tube *A* (Fig. 362) in the jar of water *B*, and at the same time hold a vibrating tuning fork over the tube. We shall find a position where the sound of the fork is reënforced by the sound of the air column and seems loudest.

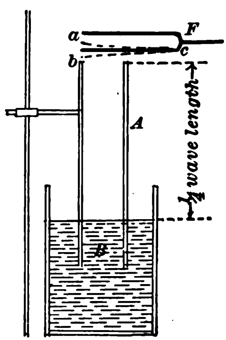


FIG. 363. — The cause of resonance

This reënforcement or intensification of sound by a resonator is due to the unison of direct and reflected waves. For example, it can be shown that the length of air column used in the ex-

periment is one quarter of a wave length. This will be readily understood from the diagram (Fig. 363), where *ac* is one prong of a fork vibrating over an air column in resonance. When the prong moves *down* past its central position, it causes a condensation in the column of air, which goes to the bottom and gets back just as the fork is moving *up* past its central position. This reënforces the vibration of the fork. Since the sound traveled twice the length of the air column in the time of half a vibration of the fork, it traveled the length of the air column in the time of a quarter vibration. So the vibrating air column is a quarter of a wave length. Further experiments would show that a resonance column may be 3, 5, 7, or any odd number of quarter wave lengths.

394. Fundamentals and overtones. When a piano wire vibrates as a whole, it gives out what is called its **fundamental** note. This fundamental is the lowest note which it can give out. Its pitch depends on the length, tension, size, and material of the wire. When a wire is vibrating as a whole, it may at the same time be vibrating in segments; that is, as if it were divided in the middle. Such a secondary vibration gives an **overtone** which has twice the frequency of the fundamental and is said to be an octave higher. Figure 364 shows an instantaneous picture of a vibrating string giving both its fundamental and its first overtone. In a similar way, a string may vibrate as a whole and, at the same time, as if divided into thirds, in which case it gives its fundamental and its second overtone. Higher overtones or "harmonics" are also possible.

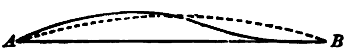


FIG. 364. — A wire emitting its fundamental and first overtone.

395. Helmholtz' experiment. Helmholtz proved that the quality of a tone is determined simply by the number and prominence of the overtones which are blended with the fundamental. To prove this, he constructed a large number

of spherical resonators (Fig. 365), each having a large opening, and also a small one adapted to the ear. A resonator of this form is especially useful because it responds easily to vibrations of *one pitch only*, and so can be used to analyze sounds. By holding each of these resonators in succession to his ear, he was able to pick out the constituents of any musical note which was being sounded, and to judge of their relative intensities. Then he reversed the process and combined these constituent overtones, reproducing the original tone. He succeeded in imitating in this way the qualities of different musical instruments and even of various vowel sounds.

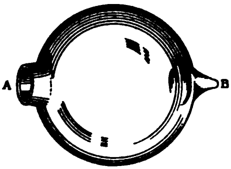


FIG. 365. — Helmholtz' resonator.

396. Koenig's manometric flames. Another method of showing that the quality of any note depends on the *form of*

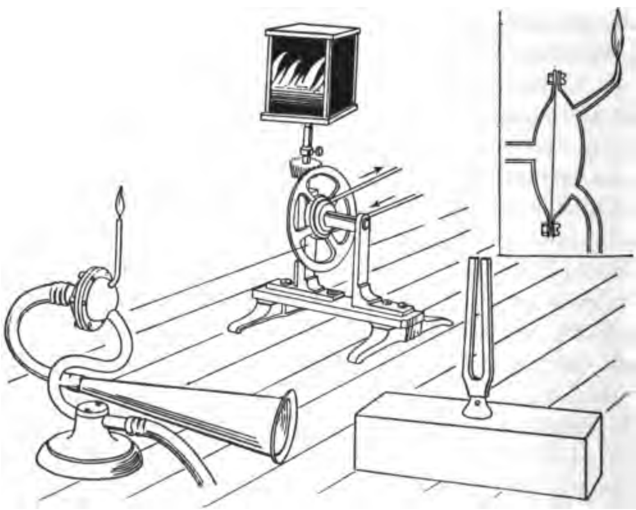


FIG. 366. — Analysis of sounds with manometric flames.

the wave was devised by a Frenchman, Koenig. This method, called **manometric flames**, has the advantage of making the phenomenon visible.

The apparatus is shown in figure 366. The essential part is a small box divided into two chambers by an elastic diaphragm, made of very thin sheet rubber or goldbeater's skin. The cavity on one side is connected with a funnel, while the cavity on the other side has two openings, one for illuminating gas to enter, and the other connected with a fine jet where the gas burns in a small flame. The vibrations of the air on one side of the diaphragm change the pressure of the gas on the other side, and cause the flame to dance up and down. When such a flame is viewed in a rotating mirror, its image is a straight band of light [Fig. 367 (top)] if the flame is still, and a serrated band [Fig. 367 (lower three curves)] when sound vibrations are striking against the diaphragm.

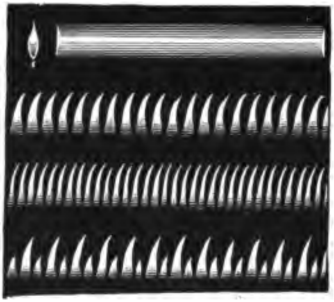


FIG. 367. — Forms shown by manometric flames.

Let us set up the apparatus as shown in figure 366, and first rotate the mirror when no note is sounded before the funnel. There will be no fluctuations in the flame as the mirror is turned. Next let a mounted tuning fork be sounded in front of the mouthpiece. Then let each of the vowels be spoken into the funnel with the same pitch and loudness. The ribbon of flame seen in the mirror is different in each case.

Manometric flames can be used to study sound vibrations of such high frequency that they are quite inaudible.

PROBLEMS

1. If two men are 1000 feet and 2500 feet from a foghorn, how many times as loud does the horn sound to one man as to the other?

2. Six seconds elapse between the firing of a gun and its echo from a cliff. If the temperature is 15° C, how far away is the cliff?

3. A tuning fork is reënforced when held over an air column 6.5 inches long. What is the wave length?

4. A tuning fork, whose normal frequency is 435, is mounted on a wooden box, which acts as a resonator. If we neglect the correction for the end, how long must the box be?

5. A whistle has a resonating column of air 1.5 inches long. Find the vibration frequency of its tone.

397. Interference of sounds. We have seen in studying resonators that two sound waves may unite so as to reënforce each other. It is also possible to make two sound waves unite so as to interfere with or destroy each other. That is, *under certain conditions the union of two sounds can produce silence.* This is the cause of the phenomenon called **beats.**

If we place two mounted tuning forks of the same pitch side by side, and strike the forks in succession with a soft mallet, we hear a smooth, even tone. But if we change the pitch of one fork by attaching a slider to one prong, and repeat the experiment, we hear a throbbing or pulsating sound. The throbs are called beats. They are due to the alternate interference and reënforcement of the sound.

If two adjoining notes of a piano or organ are struck at the same time, beats are heard, especially if the notes are in the lower part of the scale.

Beats are made use of when it is desired to tune two strings or forks to the same pitch. The forks are adjusted until no beats are heard.

398. Explanation of beats. To show how two sound waves can combine to produce no sound, let A in figure 368 represent a sound wave, and B another wave of exactly the same period, but *opposite in phase;* that is, just a half wave length behind the first. If the two impulses, which would

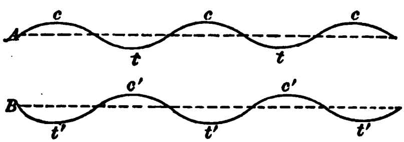


FIG. 368. — Two waves of same period but opposite phase.

generate two such waves, were applied to the air, it would not suffer any disturbance at all. This is **interference** of sound waves.

If two waves of the same period, *A* and *B*, in figure 369, are *in phase* or in step, they reënforce each other, and produce a sound of double amplitude, as shown by the dotted curve *C*. This is **reënforcement** of sound waves.

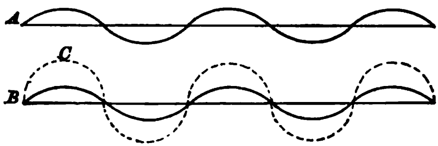


FIG. 369. — Two waves in step result in reënforcement.

Finally, if two waves of slightly different period (*A* and *B*, in figure 370) are superposed, there will be reënforcement at some points and interference at other points, as shown in the third curve *C*.

Evidently, if the waves make respectively 255 and 256 vibrations per second, there will be one reënforcement and one interference (that is, one

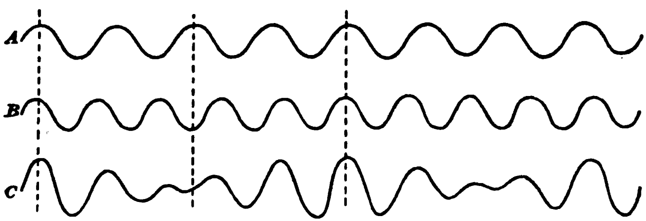


FIG. 370. — Curves to show how beats are produced.

beat) each second. In general, *the number of beats per second is equal to the difference between the frequencies of the waves.*

399. Discord and beats. Experiments show that discord is simply a matter of beats. If there are six beats or less per second, the result is unpleasant, but if there are about thirty, there is the worst possible discord. When the vibration numbers differ by as much as seventy, as do the notes

C and E, the effect is harmonious. If two musical tones with strong overtones are to be harmonious, it is essential that there shall not be an unpleasant number of beats between any of their overtones. This is the reason why the bells of chimes are struck in succession, not simultaneously.

400. The musical scale. So far we have been studying the behavior of a single train of waves in the air, and the propagation of a single musical tone; now we will consider some of the fundamental **relations** between musical tones. That is, we shall seek a scientific basis of music.

When we wish to compare two musical tones, we first consider their pitch; that is, their frequencies. Notes of the same frequency are said to be in **unison**. When two notes have frequencies as 1 to 2, the relation or interval is called an **octave**. For example, a note whose frequency is 512 is one octave *higher* than another whose frequency is 256; and one whose frequency is 128 is an octave *below* the note whose frequency is 256.

It has been found that the ear recognizes as harmonious only those pairs of notes whose frequencies are proportional to any two of the simple numbers, 1, 2, 3, 4, 5, and 6. It is still more remarkable that the ear of man has for centuries recognized that *three* notes are harmonious when their frequencies are as 4 : 5 : 6. This combination is called the **major triad.** Any combination or rapid succession of tones not characterized by simple frequency ratios produces a discord.

The **major scale** is a sequence of tones so related that the 1st, 3d, and 5th form a major triad; also the 4th, 6th, and 8th; and also the 5th, 7th, and 9th (or octave of the 2d). This is shown in the following table, where the tones of the scale are represented by the letters used in musical notation.

The arrangement of the notes of an octave on the keyboard of a piano is shown in figure 371. The white keys correspond to the notes of an octave, the black keys to intermediate notes, used in forming other scales.

TABLE OF RELATIONS BETWEEN NOTES OF AN OCTAVE

C (do)	D (re)	E (mi)	F (fa)	G (sol)	A (la)	B (si)	c (do)	d (re)
4		5		6				
			4		5		6	
				4		5		6
1	$\frac{9}{8}$	$\frac{5}{4}$	$\frac{4}{3}$	$\frac{3}{2}$	$\frac{5}{3}$	$\frac{15}{8}$	2	$\frac{9}{4}$

Any frequency or vibration number may be given to the first note C of the octave and the series built up as indicated. In fact several such pitches have been in common use as the starting point. The so-called **international pitch,** which is now almost exclusively used, takes 435 vibrations for middle A (second space on the treble cleff), and this makes middle C (the lower C on the treble cleff) 258.6. In physical laboratories C forks usually have a frequency of 256, to make the arithmetic easier.

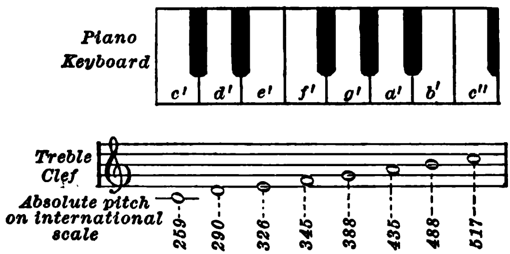


FIG. 371.—Notes of an octave on piano keyboard.

MUSICAL INSTRUMENTS

401. Piano. We are all familiar with the **piano,** or at least we have seen its keyboard, which usually has 88 keys. When we open the case, we find 88 wires of various lengths and sizes. Each key operates a little hammer which strikes a wire and thus produces a note of definite pitch. We may

C and E, the effect is harmonious. If two musical tones with strong overtones are to be harmonious, it is essential that there shall not be an unpleasant number of beats between any of their overtones. This is the reason why the bells of chimes are struck in succession, not simultaneously.

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string gives 100 vibrations per second, a pull of 16 pounds is required to raise the pitch an octave, or produce 200 vibrations per second.

(3) *The vibration frequency or pitch varies inversely as the square root of the weight per unit length of the string.* For example, the wires on the piano which give the low notes are wound with wire, to get the necessary weight.

403. Other stringed instruments. The **violin, mandolin,** and **guitar** have sets of strings tuned to give certain notes, and wooden bodies to reënforce the tones of the strings. These instruments differ from the piano in that they have but few strings, and in that their strings are set in vibration by bowing or picking instead of by striking them with a hammer. Each string is made to give a large number of notes by pressing on it at various places and so changing its length. The particular place and manner in which the string is plucked or bowed determines the overtones and thus the quality of the tone. In this way the violin may be made to give tones with a wide range not only of pitch but also of quality.

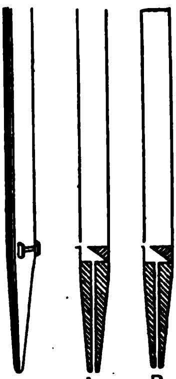

FIG. 373. — Organ pipes: (*A*) open, and (*B*) closed.

404. Wind instruments. The simplest wind instrument is the **organ pipe.** Sometimes the tube is open at the upper end and is called an *open* pipe [Fig. 373 (*A*)]; at other times the pipe is closed at the upper end and is called a *closed* pipe [Fig. 373 (*B*)].

If we blow an open pipe, the current of air strikes against a sharp edge and is set in vibration. The tube acts as a resonator. The lowest note which such a pipe gives out is the one whose wave length is twice the length of the pipe. This note is called its fundamental. If we close the end of the tube with the hand, thus making a closed pipe, we shall find that the lowest note is an octave lower, or one whose wave length is four times the length of the pipe. This is called the fundamental note of the closed pipe.

In general, then, *the length of an open pipe is one half the wave length of its fundamental, and the length of a closed pipe is one quarter of a wave length of its fundamental.*

It will be noticed that the resonance tube in the experiment in section 393 is a closed pipe upside down, the tuning-fork end corresponding to the lip end of an organ pipe.

When air is blown more violently into an organ pipe, overtones may be produced.

The **flute, clarinet, cornet,** and **trombone** are also wind instruments. In the first two, the column of air is broken up by means of holes. The opening of a hole in the tube is equivalent to cutting the tube off at the hole. In the trombone the length of the air column can be varied by sliding a portion of the tube in and out. It is also possible to vary the notes by blowing harder and so getting overtones.

In wind instruments of the **bugle** or **cornet** type, the vibration of the air is caused by the vibrating lips of the musician.

405. Vibrating membranes. One example of this sort of musical instrument is the **drum.** Another is the most wonderful musical instrument of all, the **human voice.** It is produced by the vibration of a pair of membranes on each side of the throat, called the vocal cords, and also by the vibration of the tongue and lips. By changing the muscular tension on the vocal cords one changes the pitch of his voice, and by changing the shape of the mouth, one changes the overtones, and so the quality of tone.

PROBLEMS

1. An open pipe is 4 feet long. What wave length does it give?
2. What is the length of an open pipe which gives a tone an octave above that in problem 1?
3. A siren has 50 holes. How many revolutions per minute will it have to make to produce a tone whose frequency is 435?
4. A fork making 256 vibrations per second is reënforced by a tube of hydrogen 4 feet long. What is the velocity of sound in hydrogen?

5. Find the number of vibrations of a note three octaves below a note whose frequency is 264.

6. What is the fourth overtone of a string whose fundamental tone has a frequency of 256?

7. The keyboard of a piano has 7 octaves and 2 notes. If the lowest note is A_4 (27), what is the frequency of the highest note c''''?

8. How long would an open organ pipe need to be to give the note middle A (international pitch)?

9. How many centimeters long would the closed pipe of a whistle need to be to give middle C (international pitch)?

406. The phonograph. The phonograph (Fig. 374), which was invented by Thomas Edison, is a remarkable machine

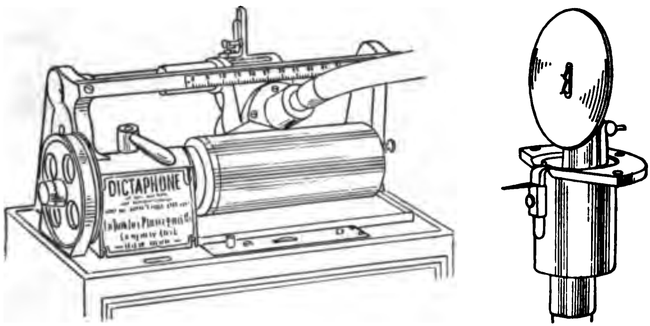

FIG. 374.—Cylinder form of phonograph and diaphragm with recording and reproducing points.

for reproducing sound, especially music and speech. When the instrument is recording sound, the waves set a diaphragm vibrating, and this makes a fine metal or sapphire point, which can move up and down, cut a spiral groove of varying depth in a wax cylinder. The bottom of this groove is a wavy line representing the condensations and rarefactions of the sound waves.

To reproduce the sound a small round-ended needle is attached to the diaphragm and follows the groove in the wax as the cylinder turns. The varying depth of the groove moves the needle up and down and thus makes the diaphragm

vibrate in such a way as to reproduce the original sounds. In the machine shown in figure 374, the sharp and the round-ended points are both mounted near the center of the same diaphragm, as shown at the right. The diaphragm can be moved forward and back a little so that only one of these points touches the cylinder at any time.

In another style of phonograph (Fig. 375) the wax is made in the form of a disk instead of a cylinder, and the

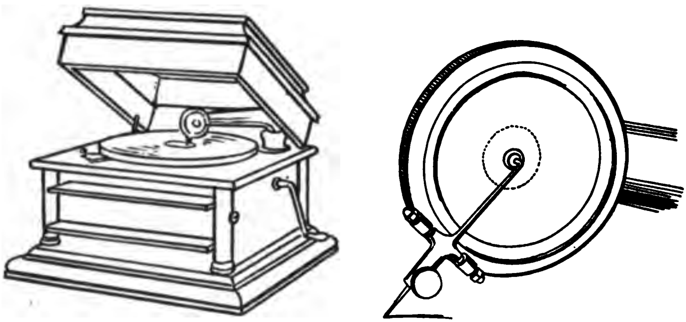

FIG. 375. — Disk form of phonograph and diaphragm.

needle point vibrates from side to side instead of up and down.

A phonograph does not reproduce the consonant sounds very distinctly, words being chiefly recognized by the vowel sounds which come out strong and clear. This is because the vowel sounds are more or less clearly defined musical tones, and produce regular vibrations, but the consonant sounds are noises produced by the mouth at the beginning and end of vowel sounds.

SUMMARY OF PRINCIPLES IN CHAPTER XX

Sound, in physics, is a vibratory motion transmitted through air or other gases, liquids, or solids.

Velocity of sound is about 1100 feet per second.

(Accurately it is 1087 ft./sec. at 0° C, and it increases about 2 ft./sec. for each degree C rise.)

Wave length = distance from crest to crest (or from condensation to condensation).

Frequency = number of waves passing given point in one second.

Velocity = frequency × wave length.

***Intensity* or loudness depends on *amplitude*.**

***Pitch* (of musical tone) depends on *frequency*.**

***Quality* (of musical tone) depends on *wave form*; *i.e.* on number and prominence of *overtones*.**

Pitch of a string (1) Rises when length is *decreased*,
(2) Rises when tension is *increased*,
(3) Is higher for small, light strings.

Length of open pipe = $\frac{1}{2}$ wave length of fundamental.

Length of closed pipe = $\frac{1}{4}$ wave length of fundamental.

QUESTIONS

1. How can the pitch of the sound from a phonograph be raised?

2. What causes a difference in the pitch of an organ pipe between a hot day in summer and a cold day in winter?

3. How can a bugler produce notes of varying pitch on an instrument of unchanging length?

4. Why is it better to bow a violin string near one end rather than in the middle?

5. Is any difference in the quality of a violin tone noticeable when the bow is moved nearer the finger board? Why?

6. How does the piano tuner go to work to tune a piano?

7. A distant band sounds much the same, except for loudness, as a band near by. What does this indicate about the velocity of sounds of different wave lengths?

8. When an electric light bulb breaks, there is a loud crash. Explain.

9. A man has two open organ pipes just alike. He saws off a little from the end of one. Explain what is heard when they are both sounded together.

10. How do the valves on a cornet operate to produce the different notes?

11. There is an old saying that "if you can count three between a flash of lightning and its thunder clap, the storm is not dangerously near." According to this how far away must the thunder cloud be for safety?

12. Explain just why the resonance experiment described in section 393 will not work if the length of the air column is half a wave length.

13. Explain how sound is produced by some form of automobile horn or signal in common use.

CHAPTER XXI

LIGHT: LAMPS AND REFLECTORS

Illumination — law of inverse squares — standard lamps and "candle power" — Bunsen photometer — "foot candles" — laws of regular reflection — plane mirrors — concave mirrors — convex mirrors — graphical construction of image — size of image — the mirror formula.

407. Problem of illumination. We have to do so much of our work and play by lamplight, that we ought to know something about illumination. Of course the first essential is to have enough light to see things distinctly. Furthermore, experience shows that we may have enough light and yet not be able to distinguish the position and shape of objects well, because the lamps are not properly distributed to cast such shadows as we are accustomed to. Then there is the very difficult problem of getting lamplight which will give colored objects the same appearance which they have in daylight. Finally, we have to protect our eyes from the glare of the modern powerful electric and gas lamps, which are likely to give us too much light in spots. Besides these purely physical aspects of the problem of illumination, we have the economic question of its cost.

408. Some optical terms. We all know that we cannot see things in a perfectly dark room and that the something which enables us to see things is **light**. There are some objects, such as the sun, the stars, and lamps, which we can see because they are **luminous**, but almost everything that we

C and E, the effect is harmonious. If two musical tones with strong overtones are to be harmonious, it is essential that there shall not be an unpleasant number of beats between any of their overtones. This is the reason why the bells of chimes are struck in succession, not simultaneously.

400. The musical scale. So far we have been studying the behavior of a single train of waves in the air, and the propagation of a single musical tone; now we will consider some of the fundamental **relations** between musical tones. That is, we shall seek a scientific basis of music.

When we wish to compare two musical tones, we first consider their pitch; that is, their frequencies. Notes of the same frequency are said to be in **unison**. When two notes have frequencies as 1 to 2, the relation or interval is called an **octave**. For example, a note whose frequency is 512 is one octave *higher* than another whose frequency is 256; and one whose frequency is 128 is an octave *below* the note whose frequency is 256.

It has been found that the ear recognizes as harmonious only those pairs of notes whose frequencies are proportional to any two of the simple numbers, 1, 2, 3, 4, 5, and 6. It is still more remarkable that the ear of man has for centuries recognized that *three* notes are harmonious when their frequencies are as 4 : 5 : 6. This combination is called the **major triad.** Any combination or rapid succession of tones not characterized by simple frequency ratios produces a discord.

The **major scale** is a sequence of tones so related that the 1st, 3d, and 5th form a major triad; also the 4th, 6th, and 8th; and also the 5th, 7th, and 9th (or octave of the 2d). This is shown in the following table, where the tones of the scale are represented by the letters used in musical notation.

The arrangement of the notes of an octave on the keyboard of a piano is shown in figure 371. The white keys correspond to the notes of an octave, the black keys to intermediate notes, used in forming other scales.

TABLE OF RELATIONS BETWEEN NOTES OF AN OCTAVE

C (do)	D (re)	E (mi)	F (fa)	G (sol)	A (la)	B (si)	c (do)	d (re)
4		5		6				
			4		5		6	
				4		5		6
1	$\frac{9}{8}$	$\frac{5}{4}$	$\frac{4}{3}$	$\frac{3}{2}$	$\frac{5}{3}$	$\frac{15}{8}$	2	$\frac{9}{4}$

Any frequency or vibration number may be given to the first note C of the octave and the series built up as indicated. In fact several such pitches have been in common use as the starting point. The so-called **international pitch**, which is now almost exclusively used, takes 435 vibrations for middle A (second space on the treble cleff), and this makes middle C (the lower C on the treble cleff) 258.6. In physical laboratories C forks usually have a frequency of 256, to make the arithmetic easier.

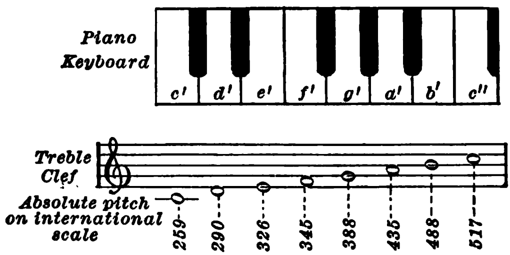

FIG. 371.—Notes of an octave on piano keyboard.

MUSICAL INSTRUMENTS

401. Piano. We are all familiar with the **piano**, or at least we have seen its keyboard, which usually has 88 keys. When we open the case, we find 88 wires of various lengths and sizes. Each key operates a little hammer which strikes a wire and thus produces a note of definite pitch. We may

also notice that the notes of lower pitch are produced by long, large wires and the notes of higher pitch by short, thin wires. Perhaps we have watched a piano tuner loosen or tighten a wire by turning with a wrench a pin at one end.

If we stretch a piece of steel wire along the table and set it vibrating, we find its tone is very weak compared to the tone of a piano. This is because the piano has a sounding board directly beneath the wires. The vibrations of the wires are transmitted through the frame to this large thin board, causing it to vibrate also. The board then sets a

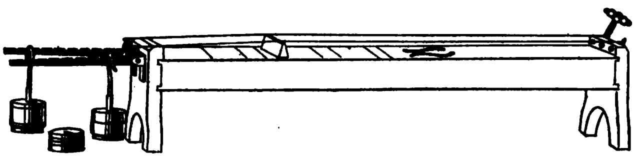

Fig. 372. — A sonometer.

larger quantity of air in vibration than the string could affect alone, and produces a louder tone.

402. Laws of vibrating strings. We may show by means of a sonometer (Fig. 372), which is simply a metal wire stretched across a long wooden box, that the pitch or frequency of a wire is raised by tightening the wire. If we introduce a movable bridge or fret, the pitch is raised. The shorter we make the wire or string, the higher is the pitch. Finally we may show that a *larger* wire of the same length and under the same tension gives a lower note.

Careful experiments of this sort have proved the following laws: —

(1) *The vibration frequency varies inversely as the length of the vibrating string.* For example, a wire under constant tension can have its pitch raised an octave by putting the movable bridge in the middle.

(2) *The vibration frequency varies directly as the square root of the tension.* For example, if a pull of 4 pounds on a

string gives 100 vibrations per second, a pull of 16 pounds is required to raise the pitch an octave, or produce 200 vibrations per second.

(3) *The vibration frequency or pitch varies inversely as the square root of the weight per unit length of the string.* For example, the wires on the piano which give the low notes are wound with wire, to get the necessary weight.

403. Other stringed instruments. The **violin, mandolin,** and **guitar** have sets of strings tuned to give certain notes, and wooden bodies to reënforce the tones of the strings. These instruments differ from the piano in that they have but few strings, and in that their strings are set in vibration by bowing or picking instead of by striking them with a hammer. Each string is made to give a large number of notes by pressing on it at various places and so changing its length. The particular place and manner in which the string is plucked or bowed determines the overtones and thus the quality of the tone. In this way the violin may be made to give tones with a wide range not only of pitch but also of quality.

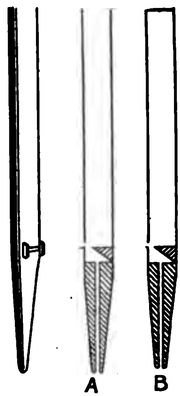

FIG. 373. — Organ pipes: (*A*) open, and (*B*) closed.

404. Wind instruments. The simplest wind instrument is the **organ pipe.** Sometimes the tube is open at the upper end and is called an *open* pipe [Fig. 373 (*A*)]; at other times the pipe is closed at the upper end and is called a *closed* pipe [Fig. 373 (*B*)].

If we blow an open pipe, the current of air strikes against a sharp edge and is set in vibration. The tube acts as a resonator. The lowest note which such a pipe gives out is the one whose wave length is twice the length of the pipe. This note is called its fundamental. If we close the end of the tube with the hand, thus making a closed pipe, we shall find that the lowest note is an octave lower, or one whose wave length is four times the length of the pipe. This is called the fundamental note of the closed pipe.

In general, then, *the length of an open pipe is one half the wave length of its fundamental, and the length of a closed pipe is one quarter of a wave length of its fundamental.*

It will be noticed that the resonance tube in the experiment in section 393 is a closed pipe upside down, the tuning-fork end corresponding to the lip end of an organ pipe.

When air is blown more violently into an organ pipe, overtones may be produced.

The **flute, clarinet, cornet,** and **trombone** are also wind instruments. In the first two, the column of air is broken up by means of holes. The opening of a hole in the tube is equivalent to cutting the tube off at the hole. In the trombone the length of the air column can be varied by sliding a portion of the tube in and out. It is also possible to vary the notes by blowing harder and so getting overtones.

In wind instruments of the **bugle** or **cornet** type, the vibration of the air is caused by the vibrating lips of the musician.

405. Vibrating membranes. One example of this sort of musical instrument is the **drum.** Another is the most wonderful musical instrument of all, the **human voice.** It is produced by the vibration of a pair of membranes on each side of the throat, called the vocal cords, and also by the vibration of the tongue and lips. By changing the muscular tension on the vocal cords one changes the pitch of his voice, and by changing the shape of the mouth, one changes the overtones, and so the quality of tone.

PROBLEMS

1. An open pipe is 4 feet long. What wave length does it give?
2. What is the length of an open pipe which gives a tone an octave above that in problem 1?
3. A siren has 50 holes. How many revolutions per minute will it have to make to produce a tone whose frequency is 435?
4. A fork making 256 vibrations per second is reënforced by a tube of hydrogen 4 feet long. What is the velocity of sound in hydrogen?

5. Find the number of vibrations of a note three octaves below a note whose frequency is 264.

6. What is the fourth overtone of a string whose fundamental tone has a frequency of 256?

7. The keyboard of a piano has 7 octaves and 2 notes. If the lowest note is A_4 (27), what is the frequency of the highest note c''''?

8. How long would an open organ pipe need to be to give the note middle A (international pitch)?

9. How many centimeters long would the closed pipe of a whistle need to be to give middle C (international pitch)?

406. The phonograph. The phonograph (Fig. 374), which was invented by Thomas Edison, is a remarkable machine

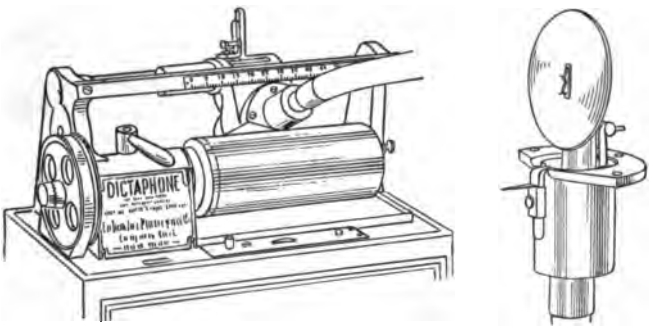

FIG. 374. — Cylinder form of phonograph and diaphragm with recording and reproducing points.

for reproducing sound, especially music and speech. When the instrument is recording sound, the waves set a diaphragm vibrating, and this makes a fine metal or sapphire point, which can move up and down, cut a spiral groove of varying depth in a wax cylinder. The bottom of this groove is a wavy line representing the condensations and rarefactions of the sound waves.

To reproduce the sound a small round-ended needle is attached to the diaphragm and follows the groove in the wax as the cylinder turns. The varying depth of the groove moves the needle up and down and thus makes the diaphragm

vibrate in such a way as to reproduce the original sounds. In the machine shown in figure 374, the sharp and the round-ended points are both mounted near the center of the same diaphragm, as shown at the right. The diaphragm can be moved forward and back a little so that only one of these points touches the cylinder at any time.

In another style of phonograph (Fig. 375) the wax is made in the form of a disk instead of a cylinder, and the

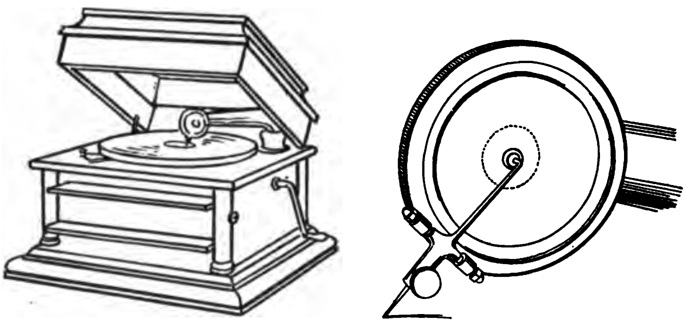

FIG. 375. — Disk form of phonograph and diaphragm.

needle point vibrates from side to side instead of up and down.

A phonograph does not reproduce the consonant sounds very distinctly, words being chiefly recognized by the vowel sounds which come out strong and clear. This is because the vowel sounds are more or less clearly defined musical tones, and produce regular vibrations, but the consonant sounds are noises produced by the mouth at the beginning and end of vowel sounds.

SUMMARY OF PRINCIPLES IN CHAPTER XX

Sound, in physics, is a vibratory motion transmitted through air or other gases, liquids, or solids.

Velocity of sound is about 1100 feet per second.

(Accurately it is 1087 ft./sec. at 0° C, and it increases about 2 ft./sec. for each degree C rise.)

Wave length = distance from crest to crest (or from condensation to condensation).

Frequency = number of waves passing given point in one second.

Velocity = frequency × wave length.

***Intensity* or loudness depends on *amplitude*.**

***Pitch* (of musical tone) depends on *frequency*.**

***Quality* (of musical tone) depends on *wave form*; *i.e.* on number and prominence of *overtones*.**

Pitch of a string (1) Rises when length is *decreased*,
(2) Rises when tension is *increased*,
(3) Is higher for small, light strings.

Length of open pipe = $\frac{1}{2}$ wave length of fundamental.

Length of closed pipe = $\frac{1}{4}$ wave length of fundamental.

QUESTIONS

1. How can the pitch of the sound from a phonograph be raised?

2. What causes a difference in the pitch of an organ pipe between a hot day in summer and a cold day in winter?

3. How can a bugler produce notes of varying pitch on an instrument of unchanging length?

4. Why is it better to bow a violin string near one end rather than in the middle?

5. Is any difference in the quality of a violin tone noticeable when the bow is moved nearer the finger board? Why?

6. How does the piano tuner go to work to tune a piano?

7. A distant band sounds much the same, except for loudness, as a band near by. What does this indicate about the velocity of sounds of different wave lengths?

8. When an electric light bulb breaks, there is a loud crash. Explain.

9. A man has two open organ pipes just alike. He saws off a little from the end of one. Explain what is heard when they are both sounded together.

10. How do the valves on a cornet operate to produce the different notes?

11. There is an old saying that "if you can count three between a flash of lightning and its thunder clap, the storm is not dangerously near." According to this how far away must the thunder cloud be for safety?

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see is visible because of the light which falls upon it and then comes from it to the eye. Such objects are **illuminated.** For example, we can see the pages of this book, if they are sufficiently illuminated, and if no obstacle is put between them and the eye. We know that light passes through some substances, like water, glass, and air, which are called **transparent,** and that practically no light gets through other substances, such as wood and iron, which are called **opaque.**

Between transparent and opaque substances there is, however, no sharp line; for example, we ordinarily think of water as transparent, and yet in the depths of the ocean utter darkness prevails. On the other hand, some opaque substances transmit light if cut in thin enough sections; for example, thin gold foil appears green when looked through. In general, light is in part turned back or **reflected** by substances, in part **transmitted,** and in part **absorbed.** An object which absorbs all the light falling upon it is called black.

409. Light advances in straight lines. Everybody knows by experience that it is impossible to see around a corner. This is because *light under ordinary circumstances advances in straight lines.*

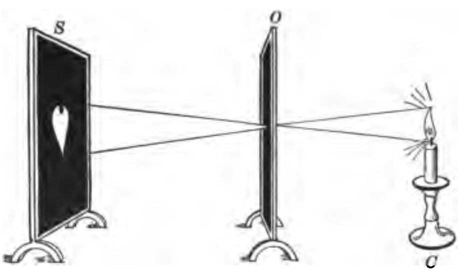


FIG. 376. — Light travels in straight lines.

If we set up a screen S and a candle C, as shown in figure 376, with an opaque screen O pierced by a pinhole in between, we see an inverted image of the flame. This shows that the light goes through the hole in straight lines. Simple "pinhole" cameras are sometimes made on this principle.

The precise measurement of angles by surveyors depends upon the fact that light comes from the distant object to the observer's instrument in straight lines.

Another consequence of this fact is the formation of a shadow when an opaque object obstructs the passage of light. The edge of the shadow is, however, a sharply defined transition between light and dark, only when the source of light is very small. For example, the shadows cast by an arc lamp are more sharply defined than those cast by a gas flame or a Welsbach mantle. This is also shown in the case of the shadow cast by the earth, as shown in figure 377. The region *A* is in the full shadow and is called the **umbra**, while in the region *BB*, on either side, the light grades off from full shadow to full illumination. This region is called the **penumbra**. When the moon happens to get wholly inside the umbra, we have what is called a *total* eclipse of the moon. When the moon is partly in the penumbra, the eclipse is *partial*.

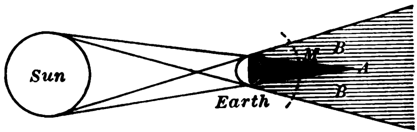


FIG. 377. — Shadow cast by the earth.

410. Intensity of illumination: law of inverse squares. It scarcely needs to be stated that a book is more brilliantly illuminated when it is held near a lamp than when it is held far from the same lamp. In other words, the intensity of illumination, that is, the amount of light falling on a unit area, decreases when the distance increases.

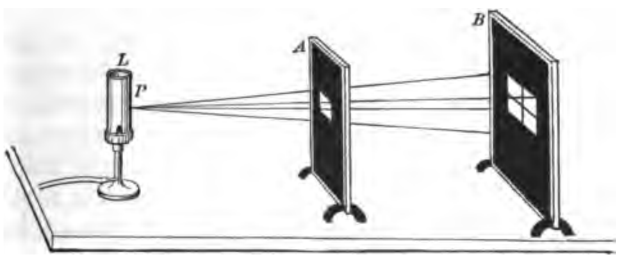


FIG. 378. — Intensity decreases as the square of the distance.

Let a sheet of metal which has in it a small pinhole *P* (Fig. 378) be set up in front of a flame, so that the source of light may be considered

a point. Then, *one foot* away, let us put a piece of cardboard A which has a hole in it one inch square. At a distance of *two feet* from the pinhole, we will put a screen B. It is evident that the light which passes through the inch hole in A is spread at B over a 2-inch square; that is, over 4 square inches. If we move the screen B so that it is 3 feet from P, the light which passes through the inch hole at A is spread over a 3-inch square; that is, over 9 square inches. The areas of these squares increase as the square of the distance. But the amount of light falling on each total area is the same. Therefore the amount on each square inch decreases as the square of the distance.

Intensity of illumination (like the intensity of sound and for the same reason) varies inversely as the square of the distance.

This law assumes that the source of light is a point, and that the surface is placed at right angles to the rays of light.

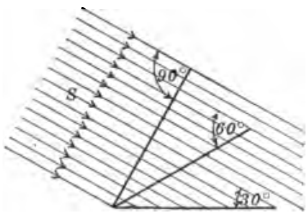


FIG. 379. — Surface not at right angles to the light.

In all practical cases, however, the source of light is a surface or region, every point of which is giving light, and in such cases this law is only approximately true. When the receiving surface is inclined (Fig. 379), it does not receive as much light per square inch as when held at right angles, and allowance has to be made for this fact.

411. Illuminating power of a lamp. In computing the amount of light received on a given area we have to consider not only the distance from the source, but also the **illuminating power** of the lamp itself. A room, for example, is much more brilliantly illuminated by a modern electric or gas lamp than by a kerosene lamp. Since there are now many different forms of lamps on the market, and every householder has to buy some kind of lamp, it is highly important that we have some way of measuring the illuminating power of a lamp. To do this we must have a **standard lamp** and some instrument for the comparison of lamps, that is, a **photometer**.

412. The standard lamp. Although many standard lamps have been proposed, none are altogether satisfactory. The oldest standard lamp, which is still used in calculation, but seldom in actual practice, is the English standard candle, which is a sperm candle made according to certain specifications. The illuminating power of a horizontal beam from this candle is called a **candle power.**

The present value of the candle power as used in the United States is that established by a set of standard incandescent lamps maintained at the Bureau of Standards in Washington, D.C. This unit of intensity is called the **international candle,** and has been accepted by England and France. In Germany the legal unit of intensity is the **Hefner**, which is equal to 0.9 international candles.

In testing gas, sperm candles are still used in routine work, although the intensity of so-called standard candles may vary by as much as 5 per cent. For more accurate work, the pentane lamp is coming into use. The Harcourt form of this lamp burns a mixture of air and pentane vapor and has an intensity of 10 candles.

The ordinary open gas flame consumes from 5 cubic feet of gas per hour upward and gives from 15 to 25 candle power. In Massachusetts the legal standard for gas is that it shall give 15 candle power in a burner consuming 5 cubic feet an hour. The gas tested by the state in 1911 averaged 18.42 c. p. Welsbach lamps consume only about 3 cubic feet of gas per hour and give from 50 to 100 candle power.

413. Bunsen photometer. This is an instrument for comparing the illuminating power of a beam from a given lamp with the illuminating power of a horizontal beam from a standard lamp. This "grease-spot" photometer was invented by the great German chemist, Robert Bunsen. It consists essentially of a white paper screen with a translucent spot in the center, which transmits light freely. The screen is placed between the lamps to be compared, so that one side is

lighted by one lamp and one by the other. If the screen is lighted more on one side, that side appears bright with a dark spot in the center, while the other side is darker with a bright spot in the center. If the two sides are equally illuminated, the spot disappears, or at least looks equally bright on each side. The arrangement of the Bunsen photometer

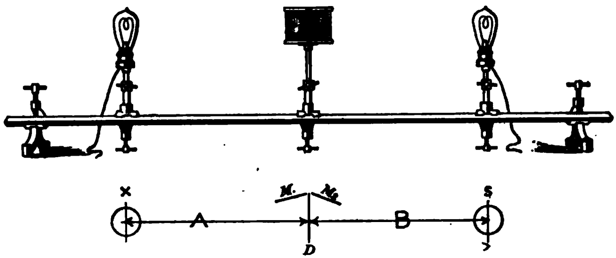


FIG. 380. — Bunsen photometer.

is shown in figure 380. The grease-spot screen is inclosed in a box, shown in figure 381, which is open at the ends A and B toward the lamps to be compared. The eye is held in front at E. Two mirrors, m_1 and m_2, are placed on either side of the screen, as indicated in the figure, so that the two sides of the screen can be seen at the same time.

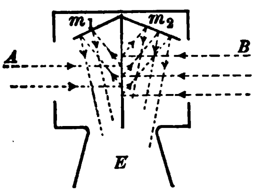


FIG. 381.— Bunsen light box with screen.

414. Use of Bunsen photometer. The photometer must be used in a dark room or else in light-tight box. The lamp X to be tested is placed at one end of the photometer bar and the standard lamp S at the opposite end. The screen is then moved back and forth until a position is found where it is equally illuminated on both sides, and the distances A and B are measured.

It is evident that if the distances A and B are equal, the candle powers of the two lamps are the same. If the distances are not equal, *the lamp which is farther from the screen*

has the greater candle power. Furthermore, since the intensity of illumination decreases as the square of the distance, *the candle powers of the two lamps are directly proportional to the squares of their distances from the screen.*

For example, let 16 = candle power of lamp S,
and X = candle power of lamp X.
Let 80 cm. = distance of screen from lamp S,
and 100 cm. = distance of screen from lamp X.
Then
$$\frac{X}{16} = \frac{(100)^2}{(80)^2};$$
so $X = 25$ candle power.

415. Distribution of light. No lamp gives light uniformly in all directions. Thus in the ordinary kerosene lamp the burner and oil reservoir cut off the light which would be radiated downward from the flame, and if the flame is broad and thin, it will give more light broadside on than edgewise. Similarly an incandescent lamp gives different intensities in different directions because of the shape of the filament.

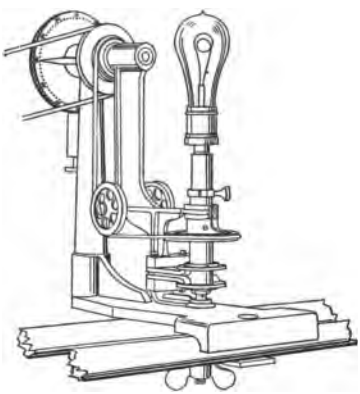


FIG. 382.—Apparatus for turning lamp to be tested.

Since an incandescent lamp can be easily turned in any position (Fig. 382), it is not difficult, with the Bunsen photometer, to measure its candle power in various positions. If the candle power is measured for several points in a horizontal plane, and the results of the tests averaged up, the result is called its **mean horizontal candle power.** Such tests show that the candle power in various directions in a horizontal plane does not vary very much. In a factory the lamp under test is rotated around a vertical axis at a speed of about 300

revolutions per minute, and the photometer reads directly the mean horizontal candle power of the lamp. A "16 candle power lamp" means a lamp of which the mean horizontal intensity is 16 candles.

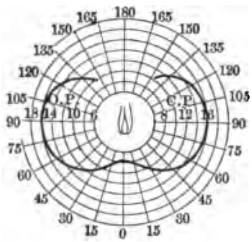


FIG. 383. — Curve to show vertical distribution.

If the lamp to be tested is tilted at various angles in a vertical plane, the results show that the lamp has very low candle power directly under the tip. The results of such tests may be best shown graphically by a diagram (Fig. 383). In this figure the intensity of the light in various directions in a vertical plane is indicated by the curve, which varies in its distance from the center of the concentric circles according to the intensity of the light. For example, the candle power directly under the tip of the bulb (0°) is a little under 8, while horizontally (90°) it is 16 candles.

When it is desirable to throw as much light as possible directly downward, some kind of a reflector or shade is used. Figure 384 shows the vertical distribution of light when the bulb is fitted with a special shade. From this curve it will be seen that the horizontal intensity is cut down to 6 candles, while the downward intensity runs over 50 candles. Such shades, made in a great variety of forms to give different desirable distributions, make it possible to work out scientifically the problem of lighting a given room or work shop efficiently.

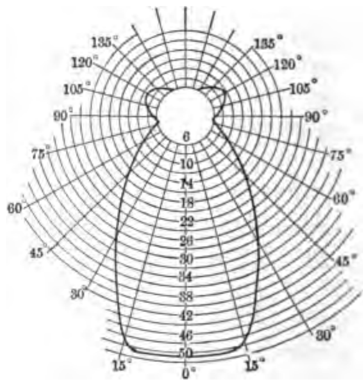


FIG. 384. — Vertical distribution, when fitted with shade.

416. Measurement of intensity of illumination. We have just seen that the unit of intensity for a source of light is the international candle. The illumination which such a standard candle throws upon a surface placed one foot away and at right angles to the rays of light is called a **foot candle.** It is the **unit of intensity of illumination.** For example, a 16 candle-power lamp would illuminate a surface placed 1 foot from it with an intensity of 16 foot candles. Again if the lamp were a 32 candle power lamp and the object were 4 feet away, the intensity of illumination would be 32 divided by $(4)^2$, or 2 foot candles.

In these examples we have assumed that there is only one source of illumination and that the surface is perpendicular to the rays of light. In practice this is almost never the case, so that the problem of computing or measuring the intensity of illumination on any given surface is very difficult. One reason for this difficulty is that we have as yet no satisfactory simple instrument for measuring intensity of illumination *directly*.

The amount of illumination needed to furnish "good light to see by" varies greatly with conditions. For example, drafting rooms, theater stages, and stores require about 4 foot candles; while churches, residences, and public corridors may need but 1 foot candle. Excessive light is as undesirable as not enough. Exposed light sources of great brilliancy (more than 5 candle power per square inch) constitute a common source of eye trouble. To avoid this, electric bulbs should be frosted and distributed in small units, or covered with shades which diffuse the light, or else concealed entirely from view, in which case the illumination is obtained by light reflected from the ceiling and walls. This indirect system of illumination gives by far the best light, especially for large rooms in public buildings, but costs more than other systems, and is to be regarded as a luxury.

PROBLEMS

1. If the page of your book is sufficiently illuminated at a distance of 8 feet from an 8 candle power lamp, how many candle power will be needed when you move 2 feet farther away?

2. If a photographic print can be made in 30 seconds when held 3 feet from a light, how long an exposure will be needed when the print is 6 feet away?

3. A 4 candle power lamp is 120 centimeters from a screen. How far away must a 16 candle power lamp be to illuminate the screen equally?

4. In measuring the candle power of a lamp, a Hefner standard lamp (0.90 candle power) is 50 centimeters from the grease spot of a Bunsen photometer, and the lamp to be tested balances it when 150 centimeters from the grease spot. How many candle power has the lamp?

5. Two lamps are 16 and 32 candle power respectively, and are 200 centimeters apart. Where between the lamps may a grease-spot photometer screen be placed for its two sides to be equally illuminated?

6. What is the illumination in foot candles on a surface 5 feet from an 80 candle power lamp?

7. The necessary illumination for reading is about 2 foot candles. How far away may a 16 candle power lamp be placed?

8. If the lamp with the special shade described in section 415 were hung above a reading table, how high should it be hung? (See curve of distribution, Fig. 384.)

9. Compare the cost of illumination with gas and electricity. A gas jet burning 5 cubic feet of gas per hour gives a flame of 18 candle power. The gas costs 85 cents per 1000 cubic feet. A 16 candle power lamp consumes 40 watts. Electricity is 10 cents per kilowatt hour.

417. Reflectors, regular and irregular. We have already said that we are able to see most objects about us by the light which they reflect to our eyes. The surface of visible objects is rough, and so the light striking the irregular surface is reflected in an irregular fashion, as shown in figure 385. This kind of reflection or turning back of the light we call **diffused reflection.** Thus the light striking a piece of paper or unvarnished wood is scattered. If, however, light strikes a flat metallic surface so carefully polished that it is very smooth,

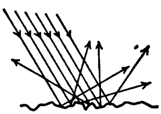


FIG. 385. — Diffused reflection from irregular surface.

the light comes to the eye as though coming directly from a distant object, instead of from the reflecting surface. This is called **regular reflection**, and is illustrated in figure 386, where *mm* is the reflecting surface or mirror. The line *OP* indicates the direction of the light falling on the mirror and *PE* indicates the direction of the reflected light.

418. Law of reflection. When light comes through a small opening, the stream of light is called a **beam**. A narrow beam may be called a **ray**.* When a beam of light comes from a very distant source, such as the sun, the rays of which it is composed are parallel, and so it is called a parallel beam.

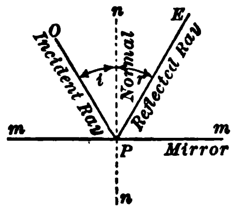


FIG. 386. — Regular reflection from smooth surface.

In figure 386, let *OP* be the direction of a parallel beam striking the mirror *mm* obliquely, and *PE* that of the reflected beam. If a line *nn*, called the **normal**, is drawn perpendicular to the reflecting surface at the point *P*, the angle between the normal and the direction *OP* of the incident beam is called the **angle of incidence**, and the angle between the normal and the direction of the reflected beam is called the **angle of reflection.**

Careful experiments have shown that, whatever the size of these angles,

I. *The incident ray, the normal, and the reflected ray lie in one plane.*

II. *The angle of incidence is equal to the angle of reflection.*

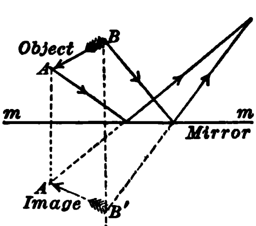


FIG. 387. — Image in a plane mirror.

419. Images in a plane mirror. We all know that if one stands in front of a plane mirror, he sees his own image and that of the objects about him, as if they were *behind the mirror*. In figure 387 we see that light coming from any

* A more accurate definition of a "ray" will be given in section 437 of the next chapter.

point A of an object is reflected by the mirror to the eye as if coming from a point A' back of the mirror. Similarly, light coming from a group of points (an object AB) seems to come from a similar group of points (the image $A'B'$) back of the mirror. The group of points from which the light *appears* to come is called the **image** of the object. A line AA' drawn from any point in the object to its corresponding point in the image is perpendicular to, and is bisected by, the mirror *mm*.

In general, *the image of an object in a plane mirror is the same size as the object, and as far behind the mirror as the object is in front.*

Indeed, such an image is so much like a real object that conjurors often make use of the illusions due to the invisibility of a well-polished mirror. It should, however, be remembered that the image is reversed from right to left, as is seen when a printed page is held in front of a mirror, so that in conjuror's tricks no letters or clock faces are allowed to be seen in mirrors.

420. Uses of plane mirrors. Good mirrors for household use are made of plate glass backed by a thin coating of silver or mercury. Only a very small fraction of the light is reflected from the front surface of the glass; the rest is reflected from the metal back.

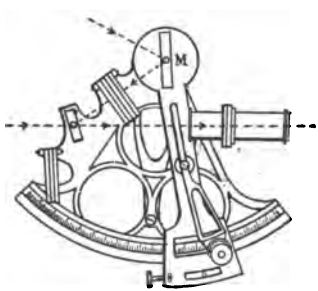


FIG. 388. — Sextant used to measure angles.

Large plate-glass mirrors are sometimes placed in the walls of public places to give the impression of spaciousness. In scientific instruments a very small mirror is often attached to a rotating part, such as the coil of a galvanometer. Such a mirror will turn a reflected beam of light through twice the angle through which the mirror itself is turned. A rotating mirror, M, is an essential part of the

sextant which the mariner uses to get the altitude of the sun. By means of a heliostat, which is simply a plane mirror turned by clockwork so as to keep up with the sun, the sun's rays may be reflected into a room, through an opening in the wall, for projection purposes.

421. Curved mirrors. A curved mirror is usually spherical; that is, it is a portion of the surface of a sphere. If it is a portion of the **outer** surface, it is called a **convex** mirror; if it is a portion of the **inner** surface, it is called a **concave** mirror. The center of the sphere, of which the curved mirror is a portion, is called the **center of curvature** (C in Fig. 389). The line CM connecting the middle of the mirror M with the center of curvature C is called the **principal axis.** Any other straight line through the center of curvature, such as CS, is called a **secondary axis.** It will be noted that any axis is perpendicular to the reflecting surface.

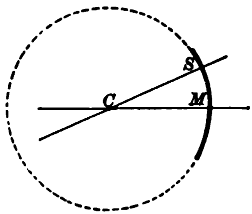


FIG. 389.—Center of a curved mirror.

422. Principal focus. When a beam of light parallel to the principal axis strikes a concave mirror, the rays are so reflected as to pass through, or very close to, a single point (F in Fig. 390).

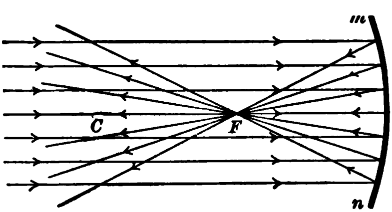


FIG. 390.—Concave mirror converges parallel rays.

This point is called the **principal focus** of the mirror. It may be defined as *that point where all rays parallel to and near the principal axis meet after reflection.*

The principal focus is located halfway between the mirror and its center of curvature.

Suppose the ray QP in figure 391, parallel to the axis AB, strikes the mirror at the point P and is reflected back in the direction PF, so as to

make the angle of incidence i equal to the angle of reflection r. Since QP and AB are parallel lines, the angle i is equal to the angle a. Therefore the angle a must be equal to the angle r, and $CF = PF$. But when P is near O, PF is nearly equal to FO, which means that F is about midway between C and O. It can be proved that the principal focus which is very close to F is exactly halfway between C and O.

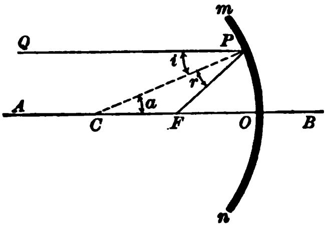


FIG. 391.—Location of principal focus.

The distance from the principal focus to the mirror is called the **focal length** of the mirror and is one half the radius of curvature.

All the rays parallel to the principal axis of a concave spherical mirror do *not* meet exactly at the same point after reflection. This failure of the rays to converge accurately at a point is called **spherical aberration**. This imperfection is slight when only a small portion of a sphere is used as a mirror. Spherical aberration in a large mirror is shown in figure 392, where it will be observed that only the central rays are reflected through the focus F, while the rays which strike the mirror near the edge are bent decidedly to the right of F.

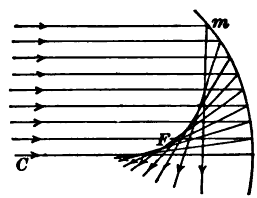


FIG. 392. — Aberration in spherical concave mirror.

It is sometimes necessary, as in the case of a searchlight, to take the divergent rays of an arc lamp and reflect them all in one direction. This can be done roughly with a concave spherical mirror, by putting the arc at the principal focus; for then the rays travel the same paths as above, but in the opposite direction. To avoid spherical aberration, however, a **parabolic mirror** (Fig. 393)

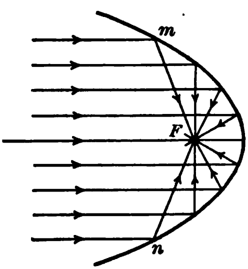


FIG. 393. — Parabolic mirror.

is generally used. These mirrors are also used in the headlights of locomotives and automobiles.

423. Applications of concave mirrors. The **ophthalmoscope** is a concave mirror with a little hole in its center. With this instrument a physician is able to reflect light from a lamp into a patient's eye, and at the same time to look through the hole into the eye thus illuminated.

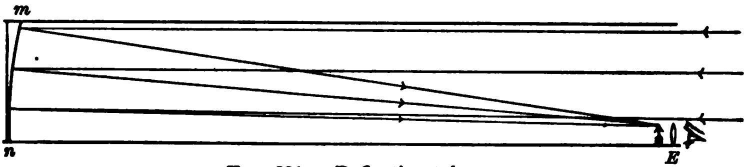


FIG. 394. — Reflecting telescope.

A certain type of telescope, called a **reflecting telescope**, consists of a long tube with a concave mirror *mn* at one end, which forms an image of a distant object. The only purpose of the tube is to support near its open end an eyepiece or magnifying glass, *E*, through which the image can be advantageously examined.

In a **compound microscope** the light from a window or lamp is concentrated upon the small object to be examined by means of a concave mirror.

We have already stated that concave mirrors are extensively used in **searchlights** and **headlights**.

424. Convex mirror. When a beam of light parallel to the principal axis strikes a convex mirror, the rays are reflected as if they came from a point behind the mirror. This is shown in figure 395, where *C* is the center of curvature and *F* is the point from which the reflected rays diverge. The point *F* is called a **virtual focus** because the rays do not

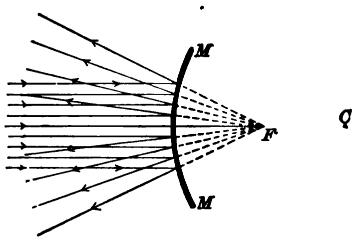


FIG. 395. — Convex mirror and virtual focus.

actually pass through it, but simply *look as if* they had come from it. In the case of the concave mirror the rays do actually pass through the point *F*, as shown by the fact that a large concave mirror of short focal length causes so great a concentration of the sun's radiant energy that paper and wood may be ignited if placed at *F*. Such a focus is a **real focus.**

425. Construction of images. It is possible to learn a great deal about the position and size of images formed by mirrors, by carefully constructing diagrams to show the paths of the rays of light.

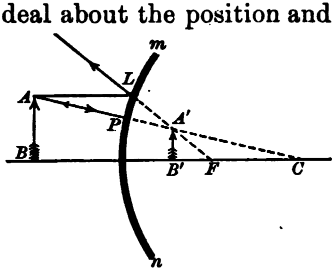


FIG. 396.—Image in a convex mirror.

Suppose *mn* in figure 396 is a convex mirror, and *AB* an object. Let us draw *AC*, a ray normal to the mirror. This ray will be reflected directly back on itself. Again let us draw *AL* parallel to the principal axis. This ray will be reflected as if it came from *F*. The image point of *A* will be where these two reflected rays cross; that is, at *A'*. Another ray from *A* that might be used in this construction is the ray through *F*. It would be reflected parallel to the axis and would also pass through *A'*.

This construction shows that the image in a **convex** mirror always seems to be **behind** the mirror and **smaller** than the object. It is **erect** and is **nearer** the mirror than the object is. *It is always a virtual image.* Thus one sees a **virtual** image of his face in a polished ball. It is always right side up and of small size.

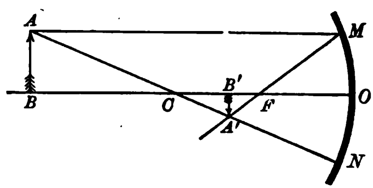


FIG. 397.—Construction of image in concave mirror.

Suppose *MON* (Fig. 397) is a concave mirror, of which *C* is the center of curvature. Let *AB* be an object which is placed beyond the center of curvature. To determine the position of the image, let us trace two rays from *A*. The point *A'*, where they intersect after reflection, is

the image of A. If AN is one such ray passing through C, it will hit the mirror perpendicularly and be reflected back along the line NC. If the other ray from A is AM, parallel to the axis, it will be reflected so as to pass through the focus F. Since B is on the axis, its image B' will also be on the axis; so that the image of the arrow AB will be the arrow $A'B'$. Another ray from A that might be used in this construction is the ray through F. It would be reflected parallel to the axis and would also pass through A'.

It will be seen that when the object is **beyond the center of curvature,** the image is **inverted** and **in front** of the mirror. Since the rays of light from A really do pass through A', the image is **real.**

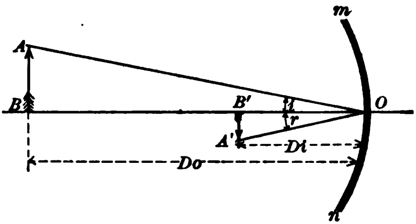


FIG. 398.—To get the size of a real image.

426. Size of a real image. Let us draw the rays AO and OA' (Fig. 398). From the law of reflection, the angle of incidence i is equal to the angle of reflection r. Therefore, the right triangles AOB and $A'OB'$ are similar, and their corresponding sides are in proportion. That is,

$$\frac{AB}{A'B'} = \frac{BO}{B'O}.$$

The size of the image is to the size of the object as the distance of the image from the mirror is to the distance of the object from the mirror.

427. Conjugate foci. We have seen that when A is the object point, the image point is at A'. But figure 397 shows that if A' is the object point, the image point is at A; for the rays will travel the same paths in the other direction. For example, if a candle were put at AB, an inverted smaller image would be formed on a screen placed at $A'B'$; also if the candle were put at $A'B'$, the image would be inverted, larger, and located at AB.

Two points, so situated that light from one is concentrated at the other, are called **conjugate foci.** For example, B and B' are two such points and therefore are conjugate foci.

428. Virtual image in a concave mirror. We have just seen that when the object is beyond the center of curvature, the image is between the principal focus F and the center of curvature C. Also when the object is between F and C, the image is beyond C. In both these cases, the image is real; that is, *the image is always real when the object is outside the principal focus F.*

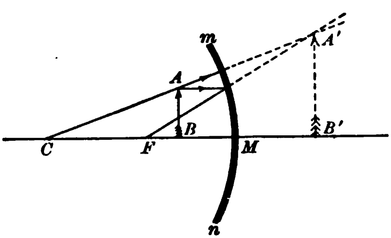


FIG. 399.—Construction of virtual image in concave mirror.

When, however, the object is placed **inside the principal focus,** that is, between F and M, as shown in figure 399, the image is **behind** the mirror, **erect, enlarged,** and **virtual.**

To show this we may, as before, trace two rays from the point A, one parallel to the axis, which is reflected through F, and the other perpendicular to the mirror, which is reflected back on itself through C. They will diverge after reflection and must be produced backward to find the point of intersection A'. The image A' is a virtual image, because the light from A does not actually pass through A'.

429. Size of a virtual image. Since every ray from A (Fig. 400) is reflected so as to seem to come from A', the ray from A to M, the middle of the mirror, will be reflected in the direction $A'MC$. Since the angles of incidence and reflection are equal,

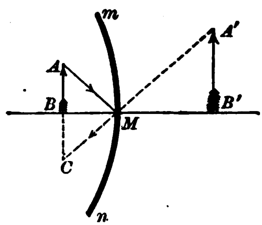


FIG. 400.—To get the size of a virtual image.

$$\text{Angle } AMB = \text{angle } BMC.$$

But $A'MB'$ and BMC are vertical angles and equal. So

$$\text{Angle } AMB = \text{angle } A'MB'.$$

Therefore the right triangles AMB and $A'MB'$ are similar, and

$$\frac{A'B'}{AB}=\frac{B'M}{BM}.$$

So in this case, as before, the size of the image is to the size of the object, as the distance of the image from the mirror is to the distance of the object from the mirror.

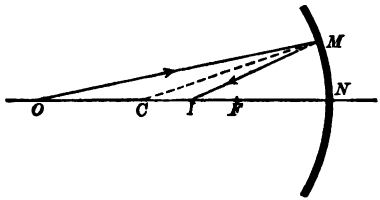


FIG. 401. — Real image in concave mirror.

430. The mirror formula. Let O be an object on the axis, OM any ray from O meeting the mirror at M. Draw the radius CM and construct the reflected ray MI, making angle OMC = angle CMI. Then I is the image of O. Since CM is the bisector of the angle OMI, it follows that

$$\frac{OM}{IM}=\frac{OC}{IC}. \tag{1}$$

Let $IN=D_I$, and $ON=D_O$. When the aperture, that is, the angle MON, is small, we have the approximate relations

$$OM=ON=D_O, \text{ and } IM=IN=D_I.$$

Now, since $FN=f$,

$$OC=ON-CN=D_O-2f,$$
$$IC=CN-IN=2f-D_I.$$

Substituting these values in the proportion (1), we have

$$\frac{D_O}{D_I}=\frac{D_O-2f}{2f-D_I},$$

and so

$$D_If+D_Of=D_OD_I.$$

Dividing by $D_O\times D_I\times f$, we have

$$\frac{1}{D_O}+\frac{1}{D_I}=\frac{1}{f},$$

where

D_O = distance of object from mirror,
D_I = distance of image from mirror,
f = focal length of mirror.

Stated in words

$$\frac{1}{\textbf{Object distance}} + \frac{1}{\textbf{Image distance}} = \frac{1}{\textbf{Focal length}}.$$

This equation gives a relation between the distance of the object, the distance of the image, and the focal length. If any two of these three quantities are known, the third can be calculated.

It can be proved that this equation holds as it stands for all cases of images either real or virtual, formed in a concave mirror. If the value of D_I comes out negative, for certain values of D_0 and f, as it will when D_0 is less than f, the meaning is that the image is behind the mirror; that is, the image is virtual. It can be shown that it holds also for convex mirrors, if the focal length f of a convex mirror is regarded as negative.

In the next chapter we shall see that the same formula holds for lenses.

PROBLEMS

1. If a ray of light strikes a plane mirror so that the angle between the ray and the mirror is 25°, what is the angle between the incident and reflected rays?

2. If the mirror in problem 1 is turned 1°, so that the angle between the incident ray and the mirror becomes 26°, through how many degrees has the reflected ray been turned?

3. An object is placed 15 inches from a concave mirror whose radius of curvature is 12 inches. How far from the mirror is the image? Is it real or virtual, erect or inverted?

4. If the object in problem 3 is 4.5 inches long, how long is the image?

5. An object is placed 12 inches from a concave mirror whose focal length is 8 inches. How far from the mirror is the image? Is it real or virtual, erect or inverted?

6. If the object in problem 5 is 2 inches long, how long is the image?

7. An arrow 1 inch long is placed 4 inches from a concave mirror whose radius of curvature is 12 inches. Find the position, length, and nature of the image.

8. If the image of a candle flame, placed 10 inches from a concave mirror, is formed distinctly on a screen 30 inches from the mirror, what is the radius of curvature?

9. How far from a concave mirror, whose focal length is 2 feet, must a man stand to see an erect image of his face twice its natural size?

10. Where must an object be placed to form, in a concave mirror whose focal length is 10 inches, a real image one half as long as the object?

SUMMARY OF PRINCIPLES IN CHAPTER XXI

Intensity of illumination varies inversely as the square of the distance.

Candle powers of lamps giving equal illumination are directly proportional to the squares of their distances from screen. (That is, lamp *farther* away is *more* powerful.)

Unit intensity of illumination, or foot candle, is illumination due to a one candle power lamp one foot away.

(Desirable intensity from 1 to 4 foot candles.)

In regular reflection: —

I. Incident, normal, and reflected rays all in one plane.

II. Angle of reflection = angle of incidence.

Plane mirror: Image always behind mirror, erect, virtual, same size as object, and at same distance from mirror as object.

Principal focus of curved mirror (either concave or convex),

***Defined* as convergence point for rays parallel to axis of mirror.**

***Located* halfway between mirror and center of curvature.**

Concave mirror: —

If object is *outside* focus, image is also outside focus, and center of curvature is between object and image. Image is inverted and real.

If object is *inside* focus, image is behind mirror, erect and virtual.

Convex mirror: Image always behind mirror, erect and virtual.

Mirror formula (holds for both concave and convex mirrors):—

$$\frac{1}{\text{Object distance}}+\frac{1}{\text{Image distance}}=\frac{1}{\text{Focal length}}.$$

For *concave* mirror, focal length is *positive*.

For *convex* mirror, focal length is *negative*.

For *real* image in front of mirror, image distance comes out *positive*.

For *virtual* image behind mirror, image distance comes out *negative*.

Size rule (holds for both concave and convex mirrors):—

$$\frac{\text{Length of image}}{\text{Length of object}}=\frac{\text{image distance (from mirror)}}{\text{object distance (from mirror)}}.$$

QUESTIONS

1. What is the difference between 16 candle power and 16 foot candles?

2. Explain how a Welsbach gas lamp consuming only 3 cubic feet of gas per hour gives over 50 candle power, while the ordinary gas jet uses 5 or more cubic feet per hour and gives only about 18 candle power.

3. If light from a very distant object, such as the sun, falls on a concave mirror, where is the image formed?

4. How does the curve of a parabola differ from the arc of a circle?

5. How does the action of a parabolic mirror differ from that of a concave spherical mirror.

6. What is the danger in too great intensity of illumination?

7. Explain how the image of a man standing in front of a plane mirror, which is tilted so as to make an angle of 45° with the floor, appears horizontal.

8. A person looking into a mirror sees a very small image of his face upside down. What kind of mirror is it?

9. Show by a diagram how a tailor arranges two mirrors so that a customer can see the back of his coat.

10. A room 20 feet square has plane mirrors on opposite walls. A man in the room holds a candle close to his head. Where should he stand so as to be as near as possible to the twice reflected image of the candle in the mirrors?

11. Why is an image in a plane mirror reversed from right to left, but not up and down?

CHAPTER XXII

LENSES AND OPTICAL INSTRUMENTS

Refraction — law of refraction — velocity of light — wave fronts — explanation of refraction — index of refraction as ratio of speeds — total reflection — prism — lens — lens formula — size rule — defects of lenses.

Camera — projecting lantern — moving pictures — eye — defects of eye — magnifying glass — microscope — telescope — erecting telescope — opera glass — prism binocular.

431. Optical instruments. The human eye is the most common and at the same time one of the most remarkable optical instruments known. Human eyes are often imperfect in various ways, and have to be "corrected," or rather aided in their work; for defective eyes themselves are seldom changed by spectacles or eyeglasses. These, too, we shall study in this chapter. Even a healthy eye has its limitations, and many optical instruments have been devised to help it to see things too far away or too small for ordinary vision. And finally, there are many devices, such as cameras, stereopticons, and moving-picture machines, that enable us to see things far away from, or long after, their actual occurrence. All these devices for enabling us to see better, farther, or at a different time are called **optical instruments**.

In all of them we find **lenses**, and in some of them also **prisms**. To understand how optical instruments work, we must first study the passage of light through lenses and prisms; that is, the **refraction of light**.

432. Refraction in water. When a stick stands obliquely in water, it appears to be broken at the surface of the water in such a way that the part under water seems to be bent

upward (Fig. 402). The bottom of a tank of water always appears to be nearer the surface than it really is. A fish appears to be higher in the water than it actually is, so that if one wishes to spear it, he must aim *under* its image. All these phenomena are due to the refraction of the light as it passes from water into air. We have said that light advances in straight lines, but this is only true in a single substance. *In general, when light goes from one substance into another of different density, it is bent or refracted at the dividing surface.*

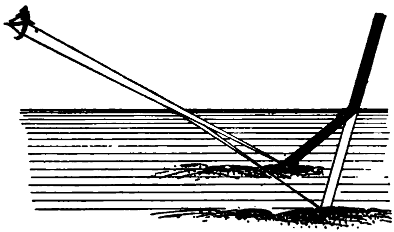


FIG. 402. — Stick partly in water appears broken.

433. Law of refraction. To measure how much a beam of light is bent in passing from water into air, we may perform the following experiment.

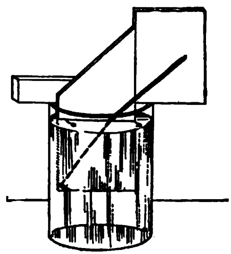


FIG. 403. — Light is bent when leaving water obliquely.

We will set a board vertically in a jar of water and fasten a wire of solder with pins along the board (Fig. 403). If we fill the jar with water, and then look down along the wire, we see that the part under water appears to be bent upward. If we bend the part that is out of water, until the whole wire seems to be straight, we have a model to show the path of the light in air and water. We may now remove the board from the water and draw the water line and the perpendicular *COD* (Fig. 404).

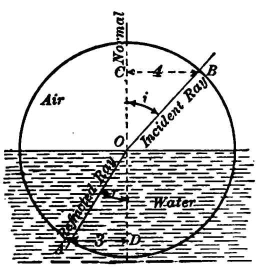


FIG. 404. — Diagram of experiment of figure 403.

From this experiment we see that a *beam of light in passing from water into air is bent away from the perpendicular.*

It might also be shown that a beam of light in passing from air into water, in the direction BO, is bent in the direction OA (Fig. 404). That is, a beam of light in passing from air into water is bent toward the perpendicular. In this case the line BO represents the **incident ray** and the line OA the **refracted ray.** The angle (COB) between the incident ray and the normal is called the **angle of incidence,** and the angle (AOD) between the refracted ray and the normal is called the **angle of refraction.** When light passes from air into water, the angle of incidence is greater than the angle of refraction.

To show the relation between the angles of incidence and refraction, we will lay off equal distances on the incident and refracted rays ($AO = BO$), and draw perpendiculars to the normal (AD and BC). We shall find that, whatever the angle of incidence, the line BC is always a definite number of times greater than AD. For example, in this case BC might be 4 inches, while AD might be 3 inches, and then the ratio BC/AD is $\frac{4}{3}$ or 1.33. This ratio is called the **index of refraction.** Experiments show that this ratio is always the same for the same two substances, no matter what the angle of incidence may be.

This ratio may also be expressed in terms of the "sines" of the angles of incidence and refraction. *Sine* is the name used in trigonometry for the ratio of the opposite side to the hypotenuse; thus the sine of the angle of incidence (i) is BC/BO and the sine of the angle of refraction (r) is AD/AO. Since $AO = BO$ by construction,

$$\frac{\text{sine of } \angle i}{\text{sine of } \angle r} = \frac{BC/BO}{AD/AO} = \frac{BC}{AD} = \text{index of refraction.}$$

434. Refraction of light by glass. We may also show that a beam of light is refracted in passing from air into glass.

Let a block of glass of semicircular shape be attached to an optical disk, as shown in figure 405. It will be seen that part of the ray is reflected by the glass as if it were a mirror, and part is refracted as it

passes into the glass. It will also be seen that the angle of incidence is equal to the angle of reflection, but is greater than the angle of refraction. We may measure the perpendicular distances from the ends of the incident and refracted rays to the normal OO, and compute the index of refraction for glass and air.

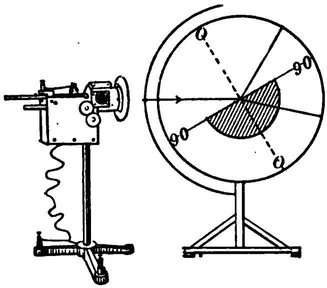


FIG. 405.—Ray is partly reflected, partly refracted.

Ordinary crown glass bends a ray of light less — that is, has a smaller index of refraction — than glass made with lead, known as flint glass. The lead glass, which is denser, has an index of refraction with respect to air of about 1.7, while that of crown glass and air is about 1.5.

In general, light is bent in passing obliquely from one substance into another, as from water to glass, diamond to air, or even from vacuum to air or from a layer of air of one density to one of another. Thus light is refracted in passing through the rising column of warm air over a stove, and things seem to shimmer or dance about. *The general rule is that the lesser angle is in the denser medium.*

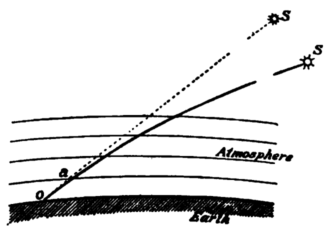


FIG. 406.—Refraction by the earth's atmosphere.

435. Some effects of refraction. An interesting case of refraction of light occurs in the atmosphere surrounding the earth. The air extends only a few miles above the surface of the earth, thinning out as it goes, and beyond is empty space. So when a ray of sunlight (Fig. 406) comes through the air obliquely, it is bent gradually toward the normal in passing from one layer to another; the result is that the eye at O sees the sun in the direction of the dotted line in the figure,

instead of in its real position. For this reason the heavenly bodies rise somewhat earlier and set somewhat later than they would if this were not the case. This makes the day some 7 or 8 minutes longer.

436. Speed of light through space. The reason for the refraction of light was not understood until the **velocity of light** in different substances had been determined. Indeed, up to 1675 it was believed that light traveled instantaneously; that is, that light consumed no time in its passage between two points. About that time Roemer, a young Danish

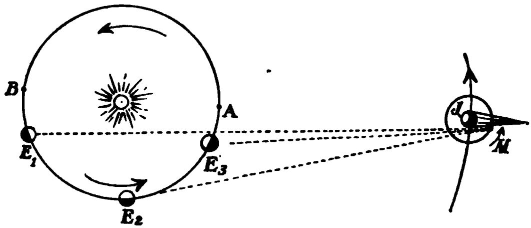


FIG. 407. — Illustrating Roemer's way of measuring speed of light.

astronomer at the Paris Observatory, was observing the moons of Jupiter. With great precision he observed just when one of the satellites M (Fig. 407) passed into the shadow cast by Jupiter, J. The beginnings of these successive eclipses of Jupiter's satellite may be thought of as signals flashed at equal intervals. When the earth is traveling away from Jupiter, the interval between signals is greater than the true interval because the light from each succeeding signal has a greater distance to travel to reach the earth. But when we are traveling toward Jupiter, the interval between signals is less than the true interval, because the light from each succeeding signal has a shorter distance to travel to reach the earth. Thus while the earth is traveling from A to B, the observed times of the eclipses are delayed more and more, and when the earth has reached B, the total

delay has amounted to 16 minutes and 36 seconds (about 1000 seconds). This means that it takes about 1000 seconds for the light to travel across the earth's orbit, a distance of 186,000,000 miles. Therefore the velocity of light is **186,000 miles per second** (300,000 kilometers per second). In recent years the velocity of light has been directly measured on the earth's surface by several methods, and while the measurements have been made with great precision, the results agree very closely with those obtained so long ago by Roemer.

This velocity is so enormous that it is not strange that the earlier experimenters could not determine it. In fact, it takes only 0.001 of a second for light to travel as far as one can see on the earth. Light travels a very little more slowly in air than in a vacuum. In denser substances, such as water and glass, light travels much more slowly.

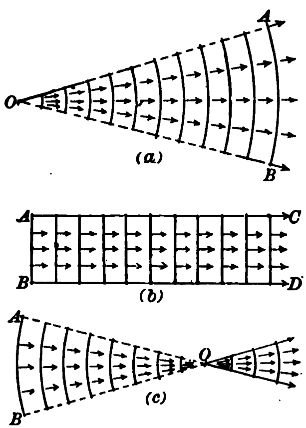


FIG. 408. — Wave fronts.

437. Light waves. Just as we think of sound as transmitted from a source through the air by a series of waves, so we think of light as transmitted through space by a series of **ether waves**. When the light comes from a point source, the "crest" or **wave front** of a wave, as it spreads in all directions with equal velocity, is spherical, and the direction of advance, being radial, is at right angles to the wave front. Figure 408 (*a*) represents such a series of expanding waves, in which the curved lines are the wave fronts and the lines of arrows indicate the direction of advance of a small section of the wave front. These lines of advance of light are what were called **rays** in the last chapter. A bundle of light rays

is a **beam.** In a "parallel beam" [Fig. 408 (*b*)] the wave fronts are plane and the rays are parallel.

By means of a lens or curved mirror a beam of light may be made to converge toward a point, called the focus. In this case the wave fronts are concave spherical surfaces which contract as they approach the focus, as shown in figure 408 (*c*).

438. Why light is refracted. When a beam of light passes from air into water, there is a change in its velocity. To see that this must cause a bending of the beam, let the parallel lines in figure 409 represent wave fronts advancing in the direction of the arrows. As soon as the edge *B* of a wave front enters the water, it begins to advance slowly, while the part *A*, which is still in the air, advances with the same speed as before. Consequently the direction of the wave front is changed into the position *CD*, and the beam is bent into a direction nearer the perpendicular *PR*.

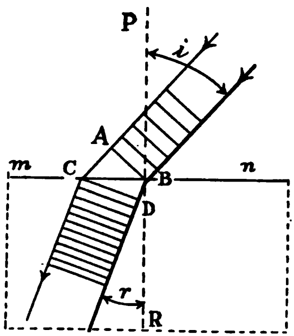


FIG. 409. Refraction of oblique waves.

This is somewhat analogous to a column of soldiers marching from a smooth, hard field into a rough, plowed field, where they are slowed up. The man at *B* hits the rough ground before the man at *A* does, and so, while *A* travels the distance *AC*, *B* has gone a shorter distance *BD*. The result is that if *B* cannot hurry, and if *A* does not slow up, the column swings around from its original direction into one nearer the perpendicular *PR*.

439. Speed of light and index of refraction. From figure 409 it will be seen that the amount which the beam of light is refracted when passing from air into water depends upon the relation between the distances *AC* and *BD*; that is, upon

the relation between the speed of light in air and its speed in water. Although it is not easy to measure the speed of light in water, yet it has been done. The speed in water has thus been proved to be about three fourths that in air. This means that the speed of light in air is 1.33 times the speed in water, which is the same number that we found for the index of refraction of water and air. It can be shown that in general

$$\text{Index of refraction} = \frac{\text{speed in air}}{\text{speed in other substance}}.$$

We may prove this as follows:

$$\text{Index of refraction} = \frac{\sin i}{\sin r} \text{ (see section 433).}$$

But i is equal to the angle ABC, and $\sin ABC = AC/BC$; also r is equal to the angle BCD, and $\sin BCD = BD/BC$.

Therefore,

$$\frac{\sin i}{\sin r} = \frac{AC/BC}{BD/BC} = \frac{AC}{BD} = \frac{\text{speed in air}}{\text{speed in water}} = \text{index of refraction.}$$

440. Sometimes no change in direction. When a stick stands vertically in water, it does not appear to be bent, because when a beam of light leaves a substance such as water *perpendicular to the surface*, it suffers no refraction. The change in velocity is, of course, just the same whether the light leaves the substance normally to the surface or obliquely, but bending or refraction occurs only when the light leaves obliquely.

441. Total reflection. We have seen in section 432 that when a beam of light passes obliquely from water or glass into air, the refracted ray is bent away from the perpendicular. For example, in figure 410 the light coming from a point O under water, in the direction oa, is refracted in the direction aa'; the ray ob is refracted along bb' and oc is refracted along cc'. As the angle in the water increases, we come finally to a ray od which is refracted along dd', and

just grazes the surface of the water. The angle which is formed between the ray *od* and the normal *NM* is called the **critical angle.** For water and air it is about 49°. If this angle is exceeded, as in the case of the ray *oe*, the ray cannot leave the water at all, but is totally reflected at *e*, just as if it had fallen on a polished metal surface, and takes the direction *ee'*.

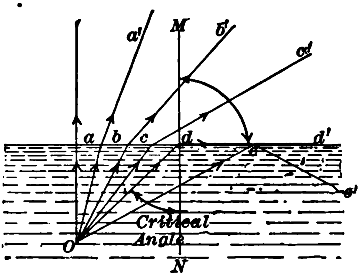


FIG. 410. — Total reflection of light by water.

The critical angle is the angle in the denser medium which must not be exceeded if the ray is to get out.

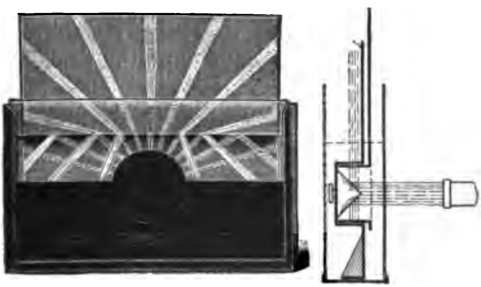


FIG. 411. — Refraction and reflection of light by water.

To illustrate total reflection, we may hold a tumbler containing water and a spoon above the eye, and look up at the surface of the water. A very bright image of the part of the spoon in the water will be seen by total reflection.

If the apparatus shown in figure 411 is available, the paths of various refracted and reflected rays, including some that are totally reflected, can be studied with great ease.

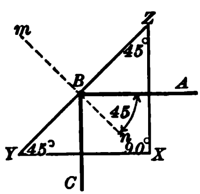


FIG. 412. — Total reflection of light by right-angle prism.

In optical instruments it is frequently necessary to have a very perfect reflector, and for this purpose a **right-angle prism** with polished sides is used. Let a ray of light *A* strike the side *XZ* of such a prism (Fig. 412) at right angles. It suffers no refraction, but passes on through the glass to *B* on the side *YZ*, where it makes an angle of 45° with the

normal *mn*. But the critical angle for crown glass is about 42°; therefore the ray AB does not emerge from the glass, but is totally reflected in the direction BC. It then strikes the face XY perpendicularly and emerges without refraction. The result is that the ray is bent 90°, as if there had been a plane mirror at YZ.

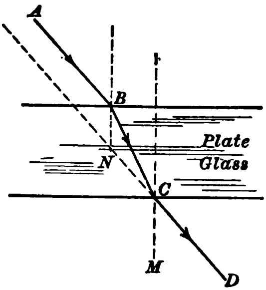


FIG. 413. — Path of ray through plate glass.

442. Refraction by plate with parallel sides. When a ray of light (AB, in figure 413) passes through a glass plate with parallel faces, such as a good window pane, it is refracted at B *towards* the normal N, and at C *away* from the normal M. The result is that the ray CD is parallel to the ray AB. Consequently when we look at any object through a glass plate, we see it slightly displaced in position, but otherwise unchanged. When the plate is thin, this change of position is too slight to attract attention.

443. Refraction by a prism. When a ray XY enters one side of a prism (ABC, in figure 414), it is bent in the direction YZ, and on emerging, it is again bent in the direction ZW. Thus the ray XO is bent out of its original course to $X'W$. The total change of direction is measured by the angle XOX', called the **angle of deviation.** Any substance which has two plane refracting surfaces inclined to each other is a **prism.**

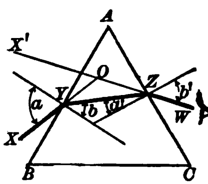


FIG. 414. — Refraction of light by a prism.

The angle A is called the **refracting angle** of the prism.

The path of a ray of light through a prism can be worked out by drawing a diagram, like figure 404, at Y and again at Z.

It should be remembered that *the beam is always bent toward the thicker part of a prism.*

PROBLEMS

(The student should have a small protractor.)

1. If the angle of incidence of a ray of light passing from air into glass is 68°, and the angle of refraction is 36°, find by construction the index of refraction.

2. If the index of refraction for air and water is 1.33, and the larger angle is 60°, find by construction the smaller angle.

3. Taking the index of refraction as 1.33, find by construction the critical angle for water.

4. If the critical angle for crown glass is 42°, find by construction the index of refraction.

5. Assuming the velocity of light in air to be about 186,000 miles per second and the index of refraction of flint glass to be 1.6, compute the velocity of light in flint glass.

6. The angles of a prism are 20°, 70°, and 90°. A ray of light enters normally the face bounded by the angles 90° and 70°. The glass has a critical angle of 42°. Prove that the ray will be twice reflected before it leaves the prism.

444. Lenses, convergent and divergent. A lens is a piece of glass, or other transparent substance, with polished spher-

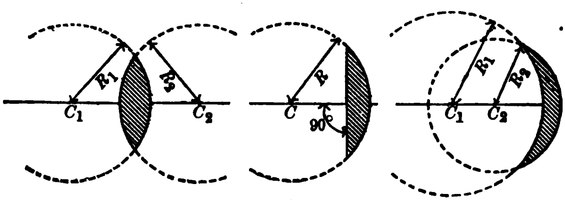


FIG. 415. — Converging lenses.

ical surfaces. A straight line drawn through the centers C_1 and C_2 (Fig. 415) of the two spherical surfaces is called the **principal axis** of the lens.

Lenses are divided into two classes, converging or "thin-edged" lenses (Fig. 415), and diverging or "thick-edged" lenses (Fig. 416). A converging lens is thinner at the edge than in the center. A common type of this class is the

double convex lens. A diverging lens is thicker at the edge than at the center. The double concave lens is a common

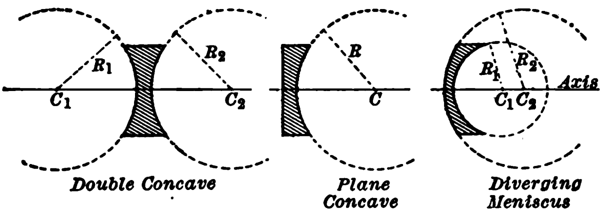


FIG. 416. — Diverging lenses.

lens of this class. It should be remembered that *when a ray of light passes through a lens, it is always bent*, just as in a prism, *towards the thicker part of the lens.*

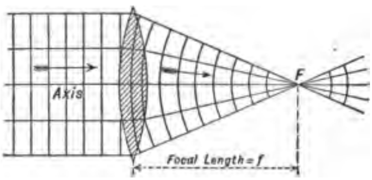


FIG. 417. — Focus of convex lens.

445. Action of converging lens. Suppose a converging lens is held so that the sunlight comes to it along its principal axis (Fig. 417). The rays of light will be so refracted as to converge at a point F on the axis. If a piece of paper is held at F, a small but very bright image of the sun is formed and the paper is quickly charred. The thicker the lens, the nearer the point F is to the lens, as shown in figure 418.

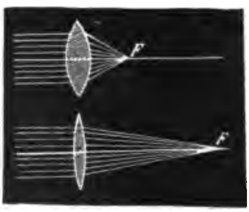


FIG. 418. — Focal length of thick and thin lenses.

The point F, where rays parallel to the principal axis converge, is called the **principal focus** of the lens. The distance from the lens to the principal focus is called the **focal length**, f, of the lens.

Since an incident ray and its corresponding refracted ray are *reversible*, it follows that a light, placed at the principal focus F, would send its rays through the lens in such a way as to come out parallel.

446. How a lens is made. The surface of a lens is shaped by grinding together the glass and an iron matrix with every possible variety of sliding motion. The glass and the matrix are thus brought automatically to an almost perfect spherical shape. The polishing is done by using finer and finer grinding materials in succession (usually powdered emery or carborundum), ending with rouge. In the later stages the matrix is lined with a layer of stiff pitch with cross grooves cut in its surfaces to hold the rouge.

447. Conjugate foci. When the light from an object *O* on the principal axis passes through a double convex lens, the rays, after leaving the glass, converge at a point *I*. Two such points, *O* and *I*, are called **conjugate foci**, for if the object were placed at *I*, the image would be at *O*. If the point *O* is not on the principal axis, the line joining *O* and *I* passes through the center of the lens, called its **optical center**, and the line is called a **secondary axis**.

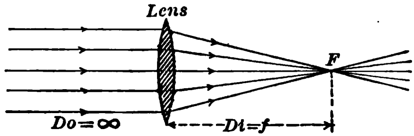


FIG. 419. — Image of distant object is at *F*.

When the lens is thin, the same formula holds as was used for mirrors (section 430).

$$\frac{1}{D_O}+\frac{1}{D_I}=\frac{1}{f},$$

where D_O = distance of object from lens,
D_I = distance of image from lens,
f = focal length of lens.

448. Discussion of the lens formula. If the object is so far away that the rays from any point of it to different parts of the lens are practically parallel, the image is formed at *F*; for D_O is very large, and so $\frac{1}{D_O}$ is nearly zero; this leads to $D_I = f$, as shown in figure 419.

If the object is brought nearer the lens, the image moves farther away from the lens. When $D_O = 2f$, $D_I = 2f$ also, as shown in figure 420.

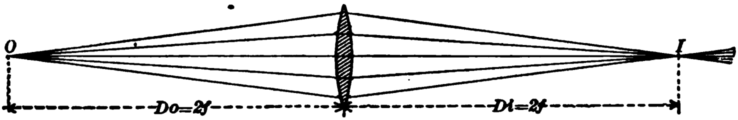


FIG. 420. — Image at same distance as object.

If the object is brought still nearer the lens, the image moves still farther away from the lens, until, when the object is at the principal focus F, the distance of the image becomes infinitely great, and the rays that go out from the lens are parallel, as shown in figure 421.

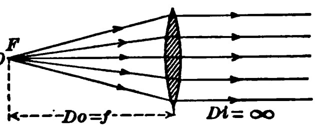


FIG. 421. — Object at F, rays emerge parallel.

If the object is brought even nearer the lens, the rays on the farther side diverge as if they came from a focus I behind the lens (Fig. 422). In this case, the formula shows that D_I is **negative.** This means that the image is **behind** the lens.

For **divergent** lenses, the **same formula** can be used, *if the focal length f is regarded as negative.*

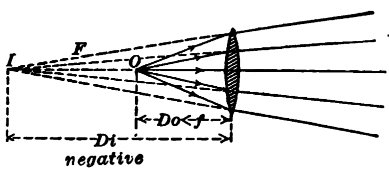


FIG. 422. — Object inside F, image virtual.

449. Images formed by lenses. The geometrical construction of images formed by lenses will indicate the size and position of these images. The method of procedure is the same as that used for spherical mirrors (section 425). If we trace two rays from any point of the object to their intersection, we have the position of the corresponding point of the image. For example, in figure 423, a ray from A parallel to the principal axis must, after refraction by the

lens, pass through the principal focus F. Another ray from A, passing through the center of lens, is undeviated. The point A' where these rays meet is the image point of A. Then from similar triangles it is readily seen that

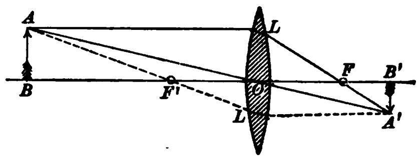


FIG. 423. — Size of real image.

$$\frac{\textbf{Length of image}}{\textbf{Length of object}} = \frac{\textbf{distance of image from lens}}{\textbf{distance of object from lens}}.$$

The ratio of the length of the image to the length of the object is called the **linear magnification.**

In figure 423 the object AB was **beyond the principal focus** of the **convex** lens, and the image $A'B'$ is **inverted, real,** and in this case smaller than the object.

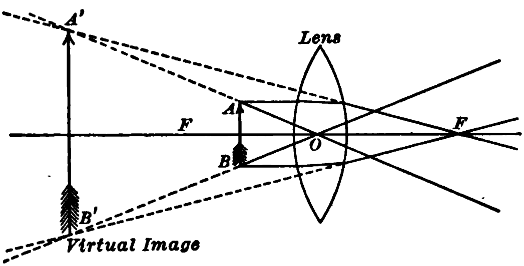


FIG. 424. — Size of virtual image.

In figure 424 the object AB is **between the principal focus** F **and the lens.** The image $A'B'$ is **erect, virtual,** and **larger,** and can only be seen by looking through the lens.

In figure 425 the lens is **concave** and the image is **erect, virtual,** and **smaller.**

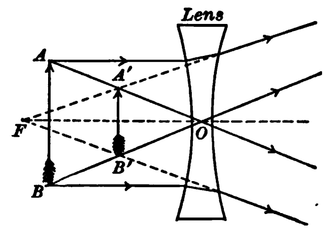


FIG. 425. — Virtual image formed by concave lens.

In all these cases it will be seen that straight lines drawn from the extremities of the object through the center of the lens pass through the extremities of the image, and therefore the diameters or lengths

of object and image are to each other as their respective distances from the center of the lens, as stated in the formula above.

450. Defects of images formed by lenses. In figure 423 it was assumed that the real image $A'B'$ was a straight line. But it will be seen that the point A of the object is at a greater distance from the center of the lens O than the point B, and, therefore, according to the lens equation, B' ought to be farther from the center of the lens than A'. In other words, the **image is curved.** This means that if a camera is equipped with a simple convex lens, and the center of the plate is sharply focused, the edges will be fuzzy, since the image does not lie in one plane. This is especially noticeable for a large object comparatively close to the lens.

In the construction of figure 423 it was assumed that all the rays coming from a point in the object are accurately refracted by the lens to one point. But as a matter of fact the rays that strike the outer portions of a lens are refracted more strongly than the rays which fall on the central portion of the lens, and so come to a focus nearer to the lens. This lack of exact concurrence is called **spherical aberration.**

The effects of spherical aberration are to make the image indistinct and to distort its shape. If the outer rays are cut out by means of a **diaphragm** or **stop**, the sharpness of the image is improved, but at the same time its brightness is diminished. In large lenses, such as those used in telescopes, the outer portions are so ground that their refracting power is diminished by the proper amount to insure distinct images.

This whole geometrical theory of lenses applies only to very thin lenses, and to cases where the light may be assumed to pass through the lens in a direction not greatly inclined to the axis of the lens. In practice, combinations of lenses are nearly always used instead of simple lenses, and these combinations are designed so that the imperfections of one lens are compensated or balanced by the imperfections of another lens.

Problems

1. A convex lens has a focal length of 16 centimeters. Find the position and nature of the images formed when objects are placed 10 meters, 50 centimeters, and 10 centimeters respectively from the lens.

2. If an object is placed 32 centimeters from the lens described in problem 1, how far is the image from the lens?

3. A lamp placed 60 centimeters from a lens forms a distinct image on a screen 20 centimeters away on the other side. Find the focal length of the lens.

Optical Instruments

451. Photographic camera. The simplest form of **camera** consists of a light-tight box (Fig. 426) with a **converging lens** at one end, so mounted as to form an image of an outside object upon a **sensitive plate.** This plate consists of a silver compound spread on a glass plate or celluloid sheet (film). The light is allowed to pass through the lens for a time which varies from a thousandth of a second up to several minutes, according to the lens, the brightness of the object to be photographed, and the "speed" of the sensitive plate. The image on the plate is not visible until the plate is placed in a mixture of chemicals called a "developer." To obviate the spherical aberration of a single lens a **diaphragm** is put in front of the lens so as to limit the size of the pencil of light. With a small opening, or "stop," we get great sharpness in the picture, but must expose it for a longer time. A "combination lens," with the diaphragm between the two lenses (Fig. 427), is used to take clear pictures

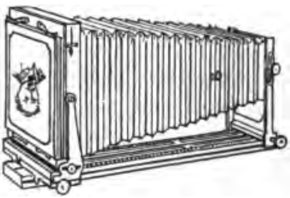


FIG. 426. — A simple camera.

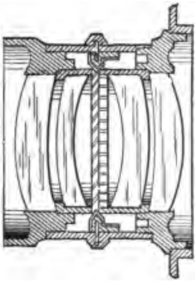


FIG. 427. — Combination lens for rapid work.

of a rapidly moving object. Since the plate on which the image is formed must be in the position which is the conjugate focus of the position occupied by the object, the camera is usually made with a bellows so that it can be "focused" on objects at varying distances.

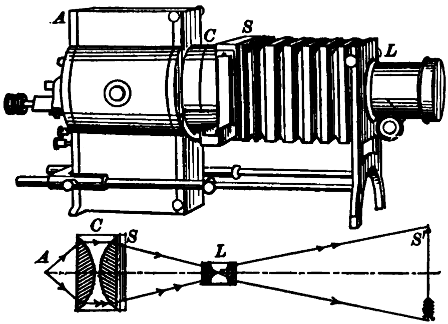


FIG. 428. — Projecting lantern.

452. Projecting lantern. The projecting lantern, or **stereopticon,** is used to throw an image of a brilliantly illuminated object or picture upon a screen. It consists essentially of a **powerful source of light,** such as an electric arc *A* (Fig. 428), the **condensing lenses** *C*, which converge the light through the slide or transparent picture *S*, and the front lens or **objective** *L*, which forms a real image of the picture on the screen *S'*. It will be noticed that the lantern is much like the camera except that the object and image have been interchanged. Since the screen is usually at a considerable distance, the slide *S* is only a little beyond the principal focus of the objective *L*. It is very important to have a powerful light source which is small in size. For this purpose electric arcs, calcium lights, acetylene lights, and electric glow lamps, in which the filament is coiled into a small space, are sometimes used. Figure 429 shows the arrangement of the lantern to project opaque pictures, such as post cards.

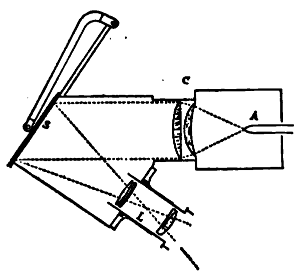


FIG. 429. — Projection of opaque objects.

The **moving-picture machine**, which is now so common, is a projecting lantern designed to show lifelike motion. A series of photographs is taken with a camera provided with a shutter which automatically opens and shuts about 12 times a second. A long narrow film moves a little while the shutter is closed, but remains stationary while it is open. Each of such a series of pictures differs slightly from the preceding one, if anything is moving in the field of the camera.

Then this series of pictures is thrown on the screen at the same rate as that at which they were taken. The sensation produced by one picture remains until the next picture appears, so that we are not aware of any interruption between the pictures.

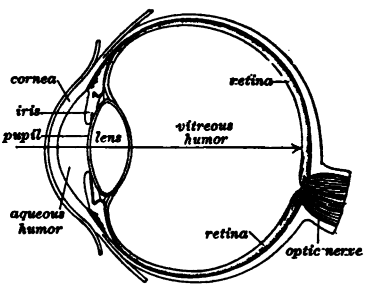


FIG. 430. — Section of the human eye.

453. The eye. The human **eye** (Fig. 430) is essentially a little camera, with a lens system in front, and a sensitive film, made of nerve fibers, at the back.

It has the great advantage over any other camera in that it can take a continual succession of pictures all on the same film, "developing" them by some unknown chemical or electrical process in the nerve fibers instantaneously, and transmitting the results equally instantaneously over a "private wire" (the optic nerve) to "headquarters" (the brain).

The structure of the eye is shown in figure 430. There is an outer horny membrane, the **cornea**, holding a watery fluid called the **aqueous humor**. There are also an adjustable diaphragm, or "stop," called the **iris**, and a **crystalline lens.** The latter is of somewhat higher index of refraction than either the aqueous humor in front or a similar fluid, **the vitre-**

ous humor, behind. At the back is the nerve layer or **retina** which acts as the sensitive film.

It should be noticed that most of the converging power of the eye comes, not in the lens, but at the front surface of the cornea. This explains why we can never see objects distinctly when swimming under water. The aqueous fluid and the water outside are so much alike that there is no longer any refraction of the light as it strikes the cornea, and the lens by itself is not powerful enough to bring the light to a sharp focus on the retina.

454. Focusing the eye. If an object is moved nearer a camera, the distance between the plate and lens must be increased, or else a lens of greater convexity, that is, of shorter focus, must be substituted, if the picture is to be sharp. Of these two possibilities, the eye chooses the second. It adapts itself to varying distances, not by moving the retina, but by changing the focal length of the lens. When the muscles of the eye are relaxed, the lens is usually of such a shape as to focus clearly on the retina objects which are at a considerable distance. When one wishes to look at near objects, a ring of muscle around the crystalline lens causes the lens to become more convex, so as to form a distinct image on the retina. It is often said that objects are seen most distinctly when held about 10 inches (25 centimeters) from the eye. This simply means that 10 inches is about as near as one can usually focus an object distinctly, and since the shortest distance gives the largest image, this is where we automatically hold an object when we want to see its details.

455. Imperfections of the eye. In the **short-sighted** eye the image of a distant object is formed in front of the retina (at *A*, in figure 431). This may be due to too great convexity in the crystalline lens, or to the oval shape of the eyeball. A person who is short-sighted must bring objects close to the eye to see them distinctly.

In the **far-sighted** eye the image of an object at an ordinary distance would be formed behind the retina (at *B*, in figure 432). This is because the crystalline lens is too flat, or the

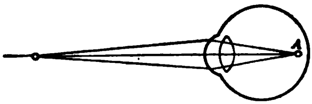


FIG. 431. — Short-sighted eye.

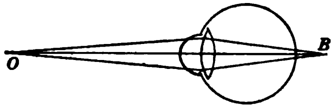


FIG. 432. — Far-sighted eye.

length of the eyeball is too short. To see distinctly, such a person must hold objects at a distance.

Spectacles with **concave** lenses are used to correct **short-sighted** eyes, and **convex** lenses are used for **far-sighted** eyes.

Another defect of the eye is **astigmatism**, which occurs when the lens of the eye, or the cornea, does not have truly spherical surfaces. The effect is that a spot of light, like a star, is seen as a short, bright line. In a case of astigmatism all the lines in such a diagram as figure 433 will not appear equally distinct. Those in one direction will be sharply defined, while those at right angles to them will appear broadened and blurred. This defect is corrected by the use of **cylindrical** lenses.

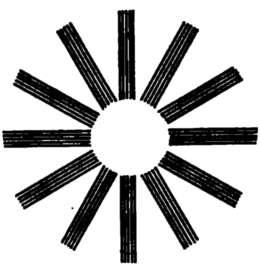


FIG. 433. — Lines to test astigmatism.

456. Apparent distance and size. The apparent size of an object depends on the size of the image formed on the retina, and consequently on the **visual angle.** From figure 434 it is evident that this angle increases as the object is brought nearer the eye.

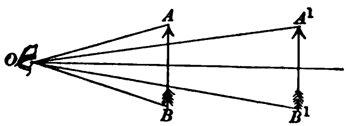


FIG. 434. — The visual angle.

For example, when we look along a railroad track, the rails seem to come nearer together as their distance from us

increases. The image of a man 100 yards away is one tenth as large as the image of the same man when he is 10 yards off. We do not actually interpret the larger image and larger visual angle as meaning a larger man, because by experience we have learned to take into account the known distance of an object in estimating its size.

Distant objects seen in clear mountain air often seem nearer than they really are. This is because we see the objects more clearly and distinguish the details more sharply; and this often leads us to think that they are smaller than they really are. The moon, on the other hand, seems bigger when near the horizon, because we can compare it with objects whose size we know. It is only by long experience that we learn to estimate the actual size and distance of objects.

457. The simple microscope or magnifying glass. We have said, in section 454, that the distance of most distinct vision is about 10 inches. If an object is placed at a greater distance than this, the image on the retina is smaller and the details of the object are not seen so distinctly. If the object is placed nearer than this, the image on the retina is blurred. When an object is examined by a **magnifying glass**, the distance between the lens and the object is made less than the focal length, and so adjusted that an erect enlarged virtual image is formed about 10 inches away (Fig. 435). The magnifying power of a simple microscope is the ratio of the size of the image to the size of the object. This is equal to the distance of the image divided by the distance of the object, that is, $10/D_o$, D_o being the distance of the object (in inches) from the lens.

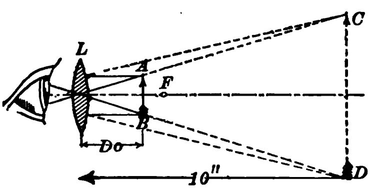
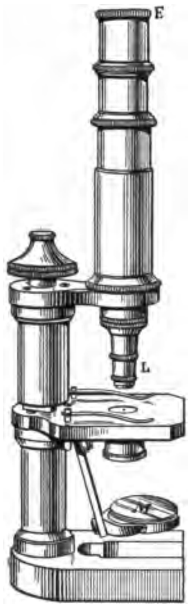


FIG. 435. — Magnifying glass.

Thus if a magnifying glass can be held 1 inch from an insect, the magnification will be 10 diameters.

458. Compound microscope. Very small objects are made visible by the **compound microscope.** It consists of *two* lenses or lens systems which are placed at the ends of a tube. The object AB is put just outside the principal focus of the smaller lens L (Fig. 436), called the **objective,** which forms an enlarged, real image CD. This real image is then ex-



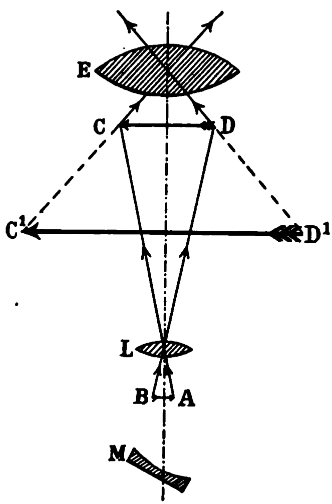


FIG. 436. — Compound microscope.

amined through the **eyepiece** E, which acts like a magnifying glass, giving a still larger virtual image at $C'D'$, about 10 inches from the eye.

The image CD is magnified as many times as its distance from the lens L is greater than the focal length of that lens. Usually the distance of CD from L is about 150 millimeters, and so, if the lens has a focal length of 5 millimeters, the

image CD is 30 times as long as the object AB. If the eyepiece still further magnifies the image 10 times, the magnifying power of the combination is 10×30, or 300 diameters. Microscopes which magnify as much as 2500 diameters are sometimes used.

We are indebted to the microscope for many of our most valuable discoveries about the structure and life of plants and animals, about the smallest living things, and about the causes of disease.

459. The telescope. The telescope enables us to see clearly objects so far away that we could not otherwise see their

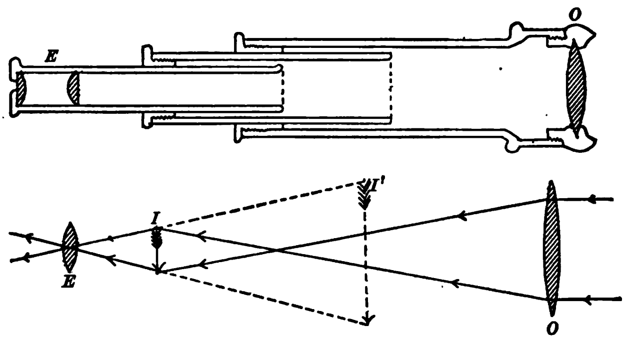


FIG. 437.—Astronomical telescope.

details. The simpler sort, called the **astronomical telescope**, consists of two lenses or lens systems, the large objective O (Fig. 437) and the eyepiece E. The inverted real image I, formed by the lens O, is much smaller than the object, but it is brought so near to the observer that it can be examined through the eyepiece E, which acts like a magnifying glass. The two lenses are mounted in an extension tube so that the eyepiece can be drawn farther from the objective when objects near at hand are to be examined. Since the magnifying glass or eyepiece does not reinvert, the observer sees things upside down, just as he does in a microscope.

It can be shown that the *magnifying power of an astronomical telescope is equal to the number of times the focal length of the eyepiece is contained in the focal length of the object glass.*

460. The erecting telescope or spyglass. This instrument (Fig. 438) is like the astronomical telescope except that an additional converging lens or lens system L is introduced between the object glass O and the eyepiece E. This lens L inverts the image I, forming another real image at I';

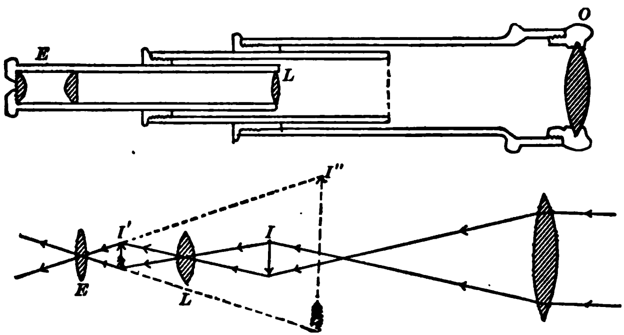


FIG. 438. — Erecting telescope or spyglass.

then this erect image I' is magnified by the eyepiece, which forms an enlarged, erect, virtual image I''. In the ordinary spyglass the eyepiece is a combination of two lenses, which act like a single magnifying glass. The introduction of the erecting lens L lengthens the telescope tube considerably.

461. Telescope used for sighting. A gun cannot be sighted with the greatest possible accuracy if its sights are pins or pointed projections. This is because it is impossible to focus the eye both on the sights and on a distant object at the same time. For example, the best that can be done with the naked eye at a distance of 100 yards is subject to error of one or two inches. Therefore many of the best long-range rifles are provided with telescopic sights. Similarly, surveyors make use of the telescope in their "transits" and

"levels." In all such cases two very fine wires or spider lines are stretched across the telescope in the plane where the image of the distant object is formed by the object glass, and the intersection of these two cross hairs is made to coincide with the image of any given point of the object. When this adjustment is made, a line drawn from the point of intersection of the cross hairs through the center of the object glass passes through the given point of the object.

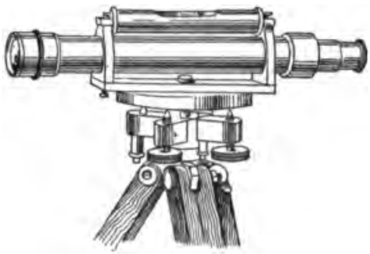


Fig. 439. — Surveyor's level.

462. The opera glass and field glass. The **opera glass** (Fig. 440) is a telescope whose **eyepiece** is a **diverging** or concave lens. Since the eyepiece has approximately the same focal length as the eye of the observer, its effect is practically to neutralize the lens of the eye. So we may consider that the object glass forms its image directly on the retina. The field of view of the opera glass is small, and so the opera glass is usually made to magnify only three or four times. But it has the advantage of being compact and gives an erect image. Galileo made a telescope on this plan which magnified about 30 diameters and enabled him to make some exceedingly important discoveries. A large-sized opera glass is usually called a **field glass.**

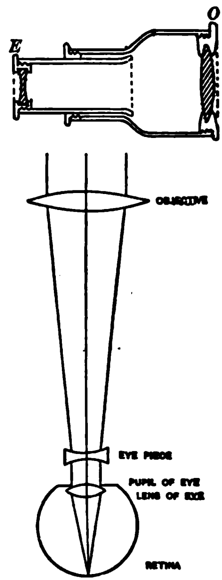


Fig. 440.—Opera glass

463. The prism field glass or binocular. An instrument, called a **binocular**, has come into use in recent years which has the wide

field of view of the spyglass and at the same time the compactness of the opera glass. This compactness is obtained by causing the light to pass back and forth between two reflecting prisms, as shown in figure 441. This device enables the focal length of the object glass to be three times as great as in the ordinary field glass for the same length of tube, and so the magnifying power is correspondingly increased.

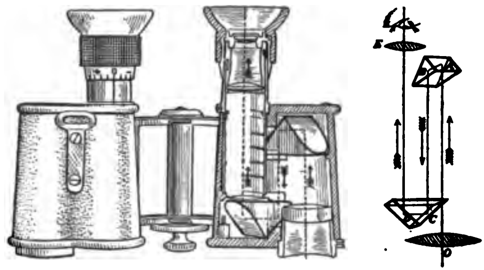


Fig. 441.—Prism binocular.

Furthermore, the reflections in the two prisms secure an erect image without using the erecting lens of the ordinary terrestrial telescope; for one double reflection tips the image right side up, and the other shifts right and left, thus restoring it completely to its natural position.

Problems

1. When a camera is focused on an automobile 100 yards away, the plate is 8 inches from the lens. How much must the distance between the lens and the plate be changed when the automobile is only 10 yards away? Must the distance be shortened or lengthened?

2. A 5 inch post card is to be projected on a screen 20 feet away so as to be 5 feet long. Find the focal length of the lens required.

3. A photographer with a "12 inch lens" wants to make a full-length picture of a 6 foot man standing 10 feet from the lens. How near the lens must the plate be placed?

4. How long a plate must be used in problem 3?

5. How near to an object must a hand magnifier of 1.2 inches focal length be held to magnify it 6 diameters?

6. A reading glass of 5 inches focal length is held 4 inches from a printed page. How much does it magnify?

7. It is necessary to project a slide 3 inches wide on a wall 40 feet

distant, so that the picture shall be 10 feet wide. What must be the focal length of the objective of the lantern?

8. In a compound microscope the objective lens L (Fig. 436) has a focal length of one inch and the object AB is 1.1 inches away. How far from the lens is the image CD? How many times is it magnified? If the eyepiece magnifies this image 20 times, what is the magnifying power of the instrument?

9. A telescope has an objective whose focal length is 30 feet, and an eyepiece whose focal length is 1 inch. How many diameters does it magnify?

10. The focal length of the great lens at the Yerkes Observatory is about 60 feet and its diameter 40 inches. The eyepiece has a focal length of 0.25 inches. Calculate its magnifying power.

SUMMARY OF PRINCIPLES IN CHAPTER XXII

Refraction occurs when light passes obliquely from one substance to another.

***Smaller* angle is always in *denser* medium.**

$$\textbf{Index of refraction} = \frac{\text{sine of larger angle}}{\text{sine of smaller angle}}, = \frac{\text{speed in rarer medium}}{\text{speed in denser medium}}.$$

Velocity of light = 186,000 miles per second,
= 3×10^{10} centimeters per second.

Critical angle is smaller angle, when larger angle is 90°.

Prism bends light toward thick edge.
Convergent (thin edged) lens bends light inward.
Divergent (thick edged) lens bends light outward.

Principal focus defined as convergence point for rays parallel to axis.

Lens formula: Holds for both converging and diverging lenses:—

$$\frac{1}{\text{Object distance}} + \frac{1}{\text{image distance}} = \frac{1}{\text{focal length}}.$$

For *convergent* lens, focal length is *positive*.

For *divergent* lens, focal length is *negative*.

For *real* image, beyond lens from object, image distance comes out *positive*.

For *virtual* image, on same side of lens as object, image distance comes out *negative*.

Size rule: Holds for both converging and diverging lenses:—

$$\frac{\textbf{Length of image}}{\textbf{Length of object}} = \frac{\textbf{image distance}}{\textbf{object distance}}.$$

QUESTIONS

1. Which people would be likely to become short-sighted early, those who live much out of doors or those who stay much indoors?

2. Compare the eye, part by part, with the camera.

3. How does a "wide-angle" lens differ from a long-focus lens?

4. What are the defects of a pinhole camera?

5. What is the difference between a refracting and a reflecting telescope?

6. Prism glass, with a section like that shown in figure 442, is often used for the upper part of shop windows and doors and for windows facing on narrow courts. Why?

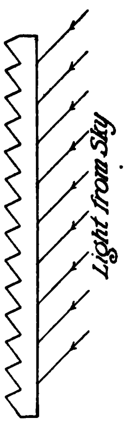


FIG. 442. — Section of plate of prism glass.

7. Why is it necessary to build powerful telescopes very wide as well as very long?

8. Why must a compound microscope be so accurately focused on the object?

9. Why is it best to have your light for writing or sewing come from over your left shoulder?

10. Explain how the wheels of moving vehicles in a moving picture sometimes seem to be rotating backwards.

11. What part do the condensing lenses play in the action of a stereopticon?

CHAPTER XXIII

SPECTRA AND COLOR

Prism spectrum — achromatic lenses — spectroscope — types of spectra — spectrum analysis — Fraunhofer lines — wave length of light — colors of objects — colors of thin films — infra-red and ultra-violet — electromagnetic theory.

464. Analysis of light by prism. If we let a beam of sunlight pass through a narrow slit into a dark room, and put a glass prism in its path (Fig. 443), the beam of light is refracted. If we put a white screen in the path of the refracted light, a band of colors is formed. In this band are red, yellow, green, blue, and violet, though there are no sharp lines of division between them.

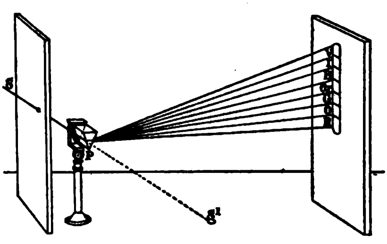


FIG. 443. — Light decomposed by prism.

This colored band, which shades off gradually from red to violet, is called a **spectrum**. This shows that the ordinary white light of the sun is complex and contains different kinds of light. The light which is refracted least, the eye recognizes as red, and that which is most refracted, as violet. It will be shown later that the physical property of light which determines this difference in refrangibility is the **wave length**.

To show that the prism itself did not produce the different colors, but simply separated various kinds of light already present in the beam of sunlight, Sir Isaac Newton placed a second prism in the spectrum, so that only violet light fell on it. He found that the violet light was again refracted, but that there was no further change in color.

He also found that when these dispersed or spread out, colored lights were brought together by a converging lens (Fig. 444), white light was the result.

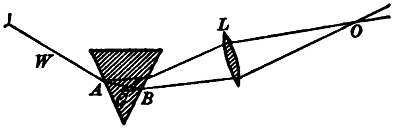


FIG. 444. — Combining spectral colors into white light.

465. Achromatic lenses. When sunlight passes through an ordinary double convex lens made of a single piece of glass, the light is refracted and converges at a point called the focus. But the light is also dispersed, just as in a prism, and the focus for red light (*R*, in figure 445) is at a greater distance from the lens than that for violet light (*V*). Such a single lens cannot give a sharp image of an object illuminated by ordinary white light, for all the lines of separation between light and dark portions of the image will be colored.

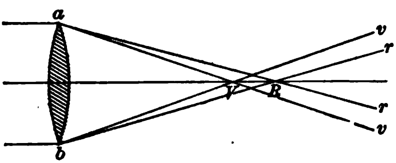


FIG. 445. — Dispersion produced by a lens.

This defect, which is known as **chromatic aberration,** may be remedied by combining a lens of crown glass with a lens of flint glass, as shown in figure 446. By carefully designing the two component lenses which are in contact, it is possible to make **achromatic lenses,** which produce the necessary refraction without dispersion. The two parts of small achromatic lenses are cemented together with Canada balsam.

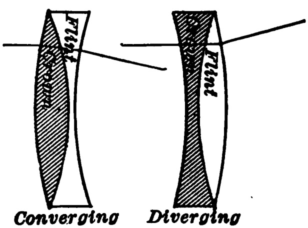
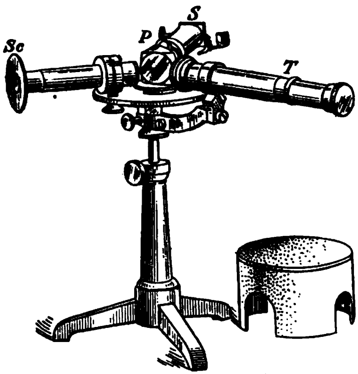


FIG. 446. — Achromatic lenses.

466. Spectroscope. In the spectrum produced by a prism the different colors overlap each other to some extent. This can be remedied by using a

spectroscope. There are four main parts in a spectroscope (Fig. 447): the **collimator**, which has a slit at one end and a convex lens at the other; a **prism**, commonly of flint glass; a **telescope**, which has an object glass and eyepiece, and a **scale tube**, which has a ruled scale at one end and a lens at the other. In the collimator the slit is at the principal focus of the lens, and so light diverging from the slit is made parallel by the lens before it reaches the prism. Here it is refracted and dispersed, each color going off as a parallel beam in its own direction. The telescope forms a sharply defined image of the spectrum. The scale tube, which is added to locate the parts of the spectrum, is so mounted that the light from the illuminated scale is reflected from the second face of the prism into the telescope along with the spectrum.



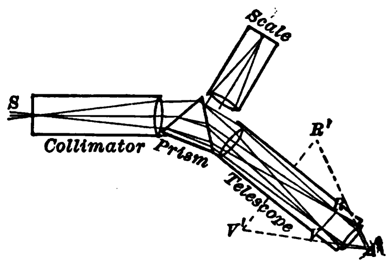


FIG. 447.—Spectroscope.

467. Kinds of spectra. The spectrum of sunlight, or **solar spectrum**, is frequently seen in summer time after a shower in the form of a rainbow. The sunlight is refracted and dispersed by the raindrops. When the solar spectrum is studied carefully with a spectroscope, it is found not to be a continuous band of colors, but to be crossed by many vertical dark lines. Since these lines were first studied by a German astronomer, Fraunhofer, they are known as **Fraunhofer lines.**

Not all sources of white light give these dark lines. For

example, an electric arc lamp, an incandescent lamp with a carbon filament, an ordinary gas flame which contains many particles of incandescent solid carbon (soot), and indeed all **incandescent solids** give **continuous spectra.**

The spectrum of an incandescent vapor or gas is quite different. It is what is called a **bright-line spectrum**, and is characteristic of the substance used.

If we dip a platinum wire or bit of asbestos into a solution of common salt (sodium chloride) and hold it in a blue Bunsen flame, we get a bright yellow flame. If we examine this flame with a spectroscope, we see a bright yellow line which occupies the position of the yellow part of the spectrum. This yellow light comes from the incandescent sodium vapor.

If we repeat the experiment with a wire dipped in a chemical, called lithium chloride, we get a red flame, which gives in the spectroscope two bands, one yellow and one red. Calcium chloride also gives two bands, green and red. (The yellow band, which is likely to be seen also, is due to sodium present as an impurity.)

468. Spectrum analysis. When the spectroscope is used to examine the spectrum of other gaseous substances, it is found that each element has its own characteristic spectrum. It may be simple as in the case of sodium, or it may be complex as in the case of iron vapor, which has more than four hundred lines. Since a very small quantity of a substance will show its characteristic spectrum lines (for example, less than one millionth of a milligram of sodium can be detected), we have a very delicate method of analyzing substances. **Spectrum analysis** was first used by the chemist Bunsen in 1859.

469. Absorption spectra. Kirchhoff (1824–1887), while a professor of physics at Heidelberg, worked conjointly with Bunsen in these investigations with the spectroscope. Kirchhoff observed that when he held an alcohol flame colored with common salt in front of the slit of the spectroscope and allowed a beam of sunlight to pass through the slit, the sodium line became especially dark and sharp, although he had expected it to be especially bright. He concluded that the

sunlight had been in part absorbed by the yellow sodium flame and that the special part had been removed which the sodium flame itself ordinarily gives out. This fact was generalized by Kirchhoff in the following law: —

A glowing gas absorbs from the rays of a hot light-source those rays which it itself sends forth.

The demonstration of Kirchhoff's law may be conveniently performed with the apparatus shown in figure 448. The source of light L is the glowing positive carbon of the electric arc, whose rays are made parallel by a lens O. Two strips of asbestos board, soaked in salt water, are

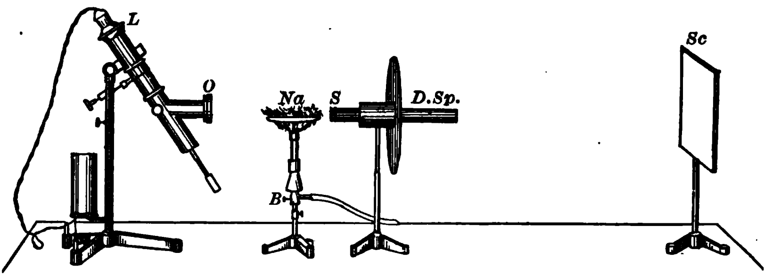


FIG. 448. — Absorption of light by sodium vapor.

heated by a wing top Bunsen burner. The light from the electric arc passes directly through the sodium flame into a "direct-vision" spectroscope which disperses the light on the screen Sc.

First we set the sodium flame burner to one side, and produce a continuous pure spectrum on the screen.

Then we bring the sodium flame into position, and we see in the yellow portion of the spectrum a dark line.

If we cover the lens O with an opaque cardboard, of course the spectrum disappears, but in the place of the dark line we now have the bright sodium line.

Finally, if we place a small white screen with a narrow slit where the dark line is located just in front of the screen Sc, the dark line on the screen Sc shows as a yellow line.

This shows that the dark absorption band is not absolutely black, but is so much less intense than the direct radiation from the arc that it appears black by contrast.

It is evident, then, that to produce *black absorption lines the absorbing vapor must be colder than the luminous source.*

470. Meaning of Fraunhofer lines. We have said in section 467 that the solar spectrum contains a large number of dark lines. Kirchhoff concluded that these dark lines were caused by the presence in the glowing solar atmosphere of those substances which themselves produce bright lines in the same positions. The core of the sun is at a very high temperature and gives forth a continuous spectrum. But this core is surrounded by a layer of gas which is cooler and absorbs those light rays which it itself would send out. On this basis he concluded that such metals as iron, magnesium, copper, zinc, and nickel exist as vapors in the solar atmosphere. After much study he found that the bright-line spectra of all the elements on the earth correspond in position to certain Fraunhofer lines, and concluded that all the elements found on the earth exist in the atmosphere of the sun. There were certain other Fraunhofer lines whose elements were not known on the earth in Kirchhoff's time. One of these new elements, helium, has since been found on the earth, and perhaps the others also will sometime be found.

Kirchhoff's explanation of the Fraunhofer lines was epoch making. Helmholtz said, "It has excited the admiration and stimulated the fancy of men as hardly any other discovery has done, because it has permitted an insight into worlds that seemed forever veiled to us."

471. The nature of light. We have said that light is considered to be a vibration of the ether. That is, light and heat are both forms of radiant energy. But we must not think that this has always been the accepted theory. To be sure, the great Dutch physicist, Huygens (1629–1695), worked out very completely the wave theory, but his rival, Sir Isaac Newton, in England, maintained the older corpuscular theory, according to which light consists of streams of very minute particles, or corpuscles, projected with enormous velocity

from all luminous bodies. Newton's reputation as a scientist was so great that his unfortunate corpuscular theory controlled scientific thought for more than a hundred years, and it was not until the beginning of the nineteenth century that the experiments of Thomas Young in England and of Fresnel in France placed the wave theory on a firm basis.

472. Different colors due to different wave lengths. It is now possible to measure directly the length of the waves of light of different colors, and to show that the waves of red light are longest and those of violet are shortest. So in the dispersion of sunlight by a *prism*, it is the *long* waves (red) which are refracted *least*, and the *short* waves (violet) which are refracted *most*. The following table gives the approximate wave lengths of some of the colors.

WAVE LENGTHS OF LIGHT

Red,	0.000068 cm.	Green,	0.000052 cm.
Orange,	0.000065 cm.	Blue,	0.000046 cm.
Yellow,	0.000058 cm.	Violet,	0.000040 cm.

473. Colors of objects. The color of any object depends (1) on the light which illuminates it, and (2) on the light it reflects or transmits to the eye.

A skein of red yarn held in the red end of the spectrum appears red. But when held in the blue end of the spectrum, it appears nearly black. Similarly a skein of blue yarn appears nearly black in all parts of the spectrum except the blue, where it has its proper color.

Another striking experiment is to illuminate an assortment of brilliantly colored worsteds or paper flowers by the light from a sodium flame. This light contains but one group of wave lengths. Those worsteds which reflect these wave lengths look bright, while those which do not reflect them look dark. They all look either yellow or dark.

Thus it appears that when a piece of paper looks white in daylight, it is because it reflects all wave lengths equally, and when a piece of cloth looks red in daylight, it is because it reflects only those long waves which produce red light. If the white paper receives only waves of red light, it appears

red, and if the red cloth receives only waves which have no red in them, it appears dark. That is, *the color of an opaque object depends on the wave length of the light it reflects.* The Cooper-Hewitt mercury vapor lamp is a very efficient electric lamp, but it cannot be used in places when colors must be distinguished, for it does not furnish waves of red light.

If we place a piece of red glass in the path of the light which is dispersed by a prism to form a spectrum, we see only the red portion of the spectrum. This shows that all the wave lengths except the long red ones have been absorbed. In a similar way a green glass lets the green light through, but greatly reduces the other parts of the spectrum. If we insert both the green and the red glasses, the spectrum almost completely vanishes.

Thus we see that *the color of a transparent object depends on the wave length of the light it transmits.* Ordinary red glass, such as photographers use for their red lanterns, transmits freely only red light, and absorbs almost completely the yellow, green, blue, and violet light, which especially affect the chemical compounds used on photographic plates.

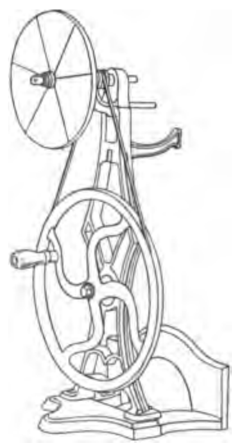


FIG. 449. — Newton's color disk.

474. Mixing colors and mixing pigments. There are other colors besides white which do not have a definite wave length. A mixture of several wave lengths may produce the same sensation as a single wave length.

Let us rotate a disk part red and part green (Fig. 449) so rapidly that the effect on the eye is the same as though the colors came to the eye simultaneously. The revolving disk appears yellow, much like the yellow of the spectrum. By mixing red and blue we get purple, which is not found in the spectrum. By mixing black with red or orange or yellow we get the various shades of brown.

The colors of the spectrum are called **pure** colors and the others **compound** colors. If yellow light is mixed with just

the right tint of blue, white light is produced. Such colors are called **complementary** colors.

Let us pulverize a piece of yellow crayon and a piece of blue crayon. If we mix the two together about half and half, the color of the resulting mixture is bright green.

This shows that while mixing yellow and blue **light** produces white, mixing yellow and blue **pigments** produces green. This is because the yellow pigment absorbs or subtracts from white light all *except yellow and green*, and the blue pigment subtracts all *except blue and green*, therefore the only color not absorbed by one pigment or the other is *green*. In other words, *in mixing pigments, the color of the mixture is that which escapes absorption by the different ingredients.*

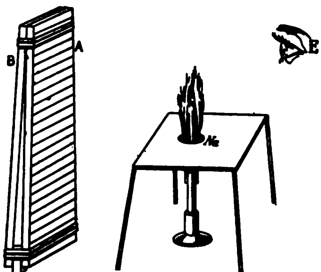


FIG. 450. — Interference of light waves.

475. Colors of thin films. The brilliant colors produced by the reflection of light from thin transparent films, like the film of a soap bubble, furnishes one of the strongest arguments for the wave theory of light.

Let us bind two pieces of plate glass A and B (Fig. 450) together with rubber bands, in such a way that they will be separated at one end by a piece of tissue paper C. If we hold the glass strips behind a sodium flame, we see in the reflected image of the yellow flame a series of horizontal fine dark lines.

To explain this effect we will draw a much-enlarged section of the glass plates with the wedge of air between. In figure 451

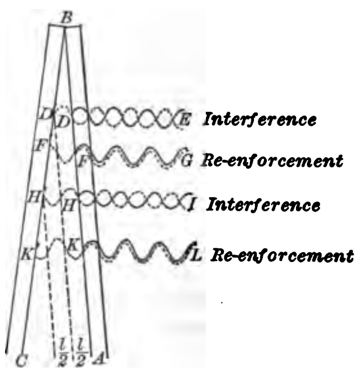


FIG. 451. — Explanation of formation of bright and dark lines.

let AB and BC be the glass plates, and let the yellow sodium light be coming from the right as a series of transverse waves which we can represent by the wavy lines. We know that this light is in part transmitted and in part reflected at each glass surface. But we are interested only in what happens at the interior faces AB and BC of the plates. Let the full line DE represent the light reflected at the point D on the surface AB, and let the dotted line $D'E$ represent the wave reflected at D' on the surface BC. If the distance from D to D' is such as to make one reflected wave just half a vibration behind the other in phase, they will neutralize each other or **interfere.** At this point we have a dark line. But at another point F the distance between the plates may be such that the wave reflected at F' coincides with or **reënforces** the wave reflected at F. At this point we see a bright yellow line. If we select any two *consecutive* dark lines, we know that the *double* path between the plates at one line must be just one wave length longer than that at the other line. This gives us a method of computing the length of a wave.

For example, we may compute the wave length of sodium light, if we know the length of the air gap, the thickness of the paper wedge, and the distance between two dark lines. Thus suppose the length of the air wedge is 100 millimeters, the thickness of the paper is 0.03 millimeters, and the distance between adjacent lines is 1 millimeter. Since the width of the wedge increases 0.03 millimeters in a distance of 100 millimeters, it increases 0.0003 millimeters in 1 millimeter, and the increase in the double path between adjacent dark lines would be 0.0006 millimeters. This is approximately the wave length of sodium light.

476. Sunlight decomposed by interference. We may substitute a soap film for the wedge-shaped air film used in the preceding experiment, and illuminate it by sunlight instead of the yellow light of the sodium flame.

Let us dip a clean wire ring into a soap solution and set it up so that the film is vertical. The water in the film will run down to the lower

edge, and the film becomes wedge-shaped. Let a beam of sunlight, or the light from a projection lantern, fall on this soap film and be reflected to a white screen. Furthermore, let a convex lens be arranged, as in figure 452, so as to produce a sharp image of the film *F* on the screen *S*. We shall see on the screen a series of horizontal bands of the various colors of the spectrum.

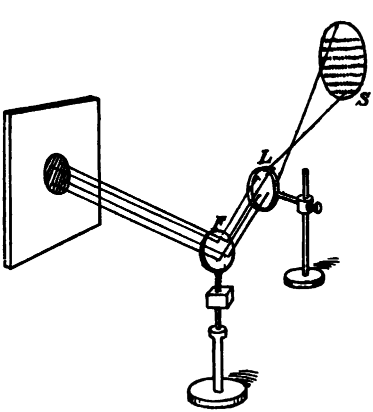


FIG. 452. — Interference of white light in soap film.

The white sunlight is composed of different colors and so of different wave lengths. The interference of the red waves takes place at one point, and that of the yellow at a different point. Where there is interference of the red waves, the complementary color, a sort of bluish-green is left; and where there is interference of the yellow waves, the color complementary to yellow, namely, blue, is produced. In this way we have a series of colored bands which are complementary to all the colors of the spectrum.

Many beautiful color effects are caused by the interference of light waves by very thin films. The colors of oil films on the surface of water, of the thin films of oxide on metals and on Venetian glass, of the feathers of the peacock and of changeable silk are due to the interference of light waves.

477. Infra-red and ultra-violet rays. In the last few years we have come to know that the sun is sending out not only the light waves which affect the optic nerve, but also other **longer** ether waves which, though invisible, yet can produce strong heating effects, and are called the **infra-red rays** (Fig. 453). We have also learned, by photographing the spectrum of the sun, that it is sending out rays too **short** to be seen, which affect a photographic plate, and are called **ultra-violet rays.**

478. Electromagnetic theory of light. As we have seen, Faraday was led to believe that his "lines of force" transmitted electricity and magnetism through some medium, probably the ether. A few years later Maxwell developed this theory of Faraday's and put it on a mathematical basis. The argument was finally clinched in 1887 by a young German, Hertz. His experiments proved that electric waves really exist, and have the same velocity as light, although

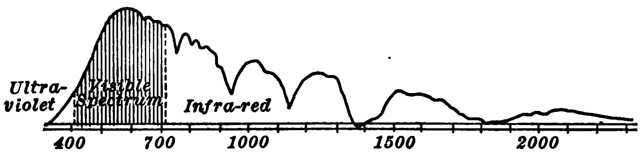


FIG. 453. — Chart of waves of varying lengths.

they are sometimes many meters long. These electromagnetic waves are reflected and refracted like light waves. Therefore, we feel sure that **light waves are electric waves.** This conception, and that of the conservation of energy, are the most remarkable achievements of physics in the nineteenth century.

SUMMARY OF PRINCIPLES IN CHAPTER XXIII

***Continuous* spectrum formed by incandescent *solids*.**
***Bright-line* spectrum formed by incandescent *gases*.**
***Dark-line* spectrum formed by incandescent solid shining through an absorbing layer of cooler gas.**

Wave length of visible spectrum ranges from about 0.000068 cm. (red) to about 0.000040 cm. (violet).

***Short* waves *most* refracted by prism.**

Color of an object depends on wave lengths reaching eye.
Colors of thin films due to disappearance of certain wave lengths by interference.

QUESTIONS

1. A clean platinum wire is held in a blue Bunsen flame and observed through a spectroscope. What sort of a spectrum would you expect to get?

2. What kind of Fraunhofer lines must one get in the light of the moon?

3. What kind of a spectrum would you get if you looked at the mantle of a Welsbach lamp through a spectroscope?

4. The complete spectrum of the sun's rays is said to consist of three parts: heat spectrum, light spectrum, and chemical spectrum. Explain the appropriateness of these terms.

5. What causes the various colored lights used in fireworks?

6. Why does a blue dress look black by the light of a kerosene lamp?

7. Why does a reddish lampshade make a room seem more cheerful at night?

8. How are colored moving pictures produced?

9. Why do they not use glass lenses in the ultra-violet microscope?

10. What sort of waves are used in wireless telegraphy?

CHAPTER XXIV

ELECTRIC WAVES: ROENTGEN RAYS

Discharge of condenser is oscillatory — electrical resonance — electric waves — detectors — wireless telegraphy.

Discharge through gases — cathode rays — Roentgen rays — radium.

ELECTRICAL WAVES

479. Discharge of Leyden jar is oscillatory. In 1842 Joseph Henry discovered that when a Leyden jar was discharged through a coil of wire surrounding a steel needle, the needle was magnetized. Not only that, but he was astonished to find that sometimes one end was made the north pole and sometimes the other, even though the jar was always charged the same way. He accounted for this fact by supposing that the discharge current kept reversing back and forth, that these oscillations gradually died away, and that the direction in which the needle was magnetized depended on which way the last perceptible oscillation happened to go. This oscillatory current is represented by the curve in figure 454.

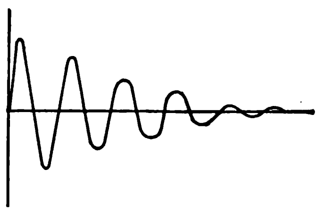


FIG. 454. — Oscillatory electric discharge.

A few years later Lord Kelvin, the great English physicist and engineer, proved mathematically that the discharge must be oscillatory. Finally, in 1859, Feddersen succeeded in photographing an electric spark by means of a rapidly rotat-

ing mirror. Figure 455 shows such a photograph. The oscillatory discharge is drawn out into a band by the rotating mirror, and thus makes a zigzag trace on the camera plate. From this experiment it is possible to calculate the time of one oscillation. It is exceedingly short, varying from one one-thousandth to one ten-millionth of a second.

480. Electrical resonance. The frequency of the oscillatory current produced by discharging a condenser depends upon the capacity of the condenser, and upon the resistance and self-induction of the circuit through which the current surges.

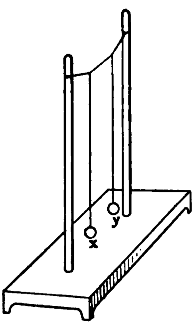


FIG. 456. — Resonance in two pendulums.

Now we have already seen, in studying sound waves, that two objects having the same frequency of vibration tend to vibrate in sympathy, and that this property of a vibrating body is called resonance. Mechanical resonance also occurs in the case of two pendulums.

Let us stretch a piece of rubber tubing between two supports and suspend two weights x and y by threads of equal length, as shown in figure 456. If we set one pendulum y swinging, the other pendulum x soon begins to swing, and the first one dies down as energy flows across to the other. This will happen only if the pendulums are of the same length and so of the same frequency. That is, *resonance* is necessary for the transfer of energy.

In a similar way, if two Leyden jar circuits have the same capacity and the same self-induction, they will have the same frequency, and one circuit will influence the other.

In figure 457 let A and B be two Leyden jars of the same size and thickness of wall. To the jar A is connected a rectangular circuit of thick wire, one end of which touches the outer coating of the jar, while the other is separated from the knob of the jar by a small spark gap. The jar B is connected to a similar circuit, except that the end CD of the rectangle can be slid back and forth, and there is no spark

Fig. 455 (in upper corner). — Oscillations of electric spark.

Fig. 466. — X-ray picture of a broken ankle, which had been called "sprained" by a doctor. Taken in a physics laboratory by the brother of the patient.

gap. Finally, let the inner coating of B be connected to its outer coating by a strip of foil cut sharply across at X.

If we place the two electrical circuits a foot apart and parallel, and send sparks across the gap of A by means of an induction coil, we find that there is a position of the slider CD such that tiny sparks appear at the gap X in the foil strip on B. When the slider is moved a short distance from this position either way, the sparks at X cease.

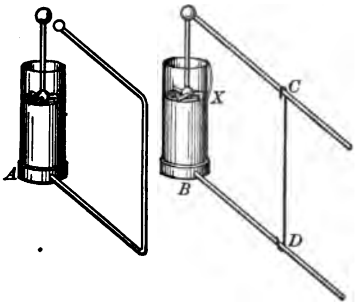


FIG. 457.—Resonance between electrical circuits.

This phenomenon is called **electrical resonance.** Although there is no connection between the two circuits, yet the energy in one circuit surges over into the other, which is in tune with it, and causes a spark there. In seeking for an explanation of this experiment, and many others, we must conclude that an oscillatory discharge or spark sends out **waves** in the surrounding ether. The ether does for the electric circuits what the rubber tubing did for the pendulums. It serves as a medium for the transfer of energy.

These electric waves were first detected and measured by Hertz, in 1888, and are therefore called **Hertzian waves.** They travel with the same velocity as light.

481. Electric wave detectors. Very sensitive means for detecting these electric waves have been invented. One means, invented by Branly and used by Marconi in his first wireless telegraphs, is called a **coherer.** It consists of a small glass tube closed at each end by metal pistons. A space of a millimeter or two between the pistons is filled with rather coarse filings of nickel and silver. When electric waves fall on this coherer, the mass of filings "coheres" or sticks together and becomes a conductor. A slight tap causes the resistance of the coherer to return to its original high value.

The **microphone**, described in section 310, is an excellent wave-detector. Another form, called a **crystal detector**, consists of a piece of silicon, or of any one of several crystalline substances, such as galena, embedded in soft metal on one side and touched on the other by a metal point. In the **electrolytic detector** a fine metal point just touches the surface of a conducting solution or electrolyte. The operation of crystal and electrolytic detectors seems to depend on some mysterious property whereby they let electricity flow through them in one direction much more easily than in the other.

482. Wireless telegraphy. Through the efforts of the Italian inventor, Marconi, and many others, electric waves are now being extensively used in wireless telegraphy.

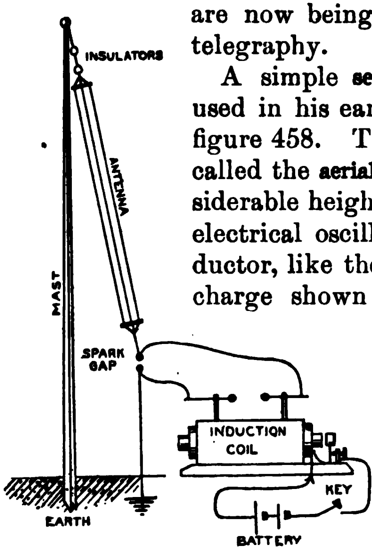


FIG. 458. — Simple sending station.

A simple **sending station**, such as Marconi used in his earliest experiments, is shown in figure 458. The essential part is a conductor called the **aerial** or **antenna**, extending to a considerable height above the ground. Powerful electrical oscillations are set up in this conductor, like the oscillations in the spark discharge shown in figure 455. These send waves out through the ether, just as a stick laid on water and shaken up and down sends out ripples over the surface of the water.

One way to set up oscillations in an aërial is to put a spark gap in it, and to send sparks across this gap by means of an induction coil fed by batteries, as in figure 458. Another way is to put a condenser in parallel with the gap in the aërial, and to fill the condenser many times a second by means of a step-up transformer fed by an alternating current.

Such a condenser will send a spark across the gap each time it fills. Another way is to put a very high frequency alternating current dynamo in the place of the spark gap in the antenna.

The simplest kind of a **receiving station** is represented in figure 459. There is an aërial like that at the sending station, except that instead of a spark gap, it contains a detector of some sort. In parallel with this detector is a telephone receiver. Every time a train of waves reaches such a receiving station, some of the energy is absorbed by the aërial, and electrical oscillations are set up in it. These cannot get through the telephone because of its self-induction, and so they have to pass through the detector. But since a crystal detector lets more electricity through one way than the other, an excess of electricity accumulates in the antenna. This excess then discharges through the telephone, and the diaphragm moves over and back once. Since this happens every time a train of waves comes in, which is many times every second, as long as the key of the sending station is closed, the telephone diaphragm is kept vibrating and emits a steady musical note. The duration of this note can be made shorter or longer by holding the sending key down a shorter or a longer time, and so the dots and dashes of the Morse code can be transmitted.

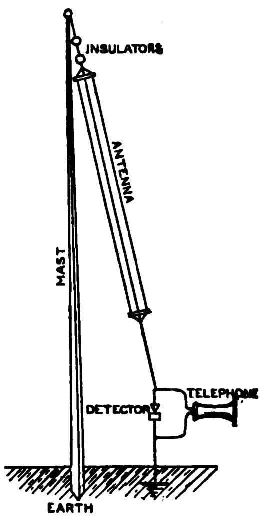


FIG. 459. — A simple receiving station.

The circuits used in commercial wireless telegraphy are much more complicated than these, because it is necessary to "tune" the sending and receiving stations accurately to the same frequency, and to make them insensitive to waves

of any other frequency, so that one pair of stations may not interfere with another. For an explanation of commercial sending and receiving stations the reader may consult any of the numerous popular or technical books on wireless telegraphy.

Wireless telegraphy is now used on all ocean steamships, so that they are in constant communication with other ships or with land stations. Timely aid has thus been called to ships in distress. Warships are kept in touch with the naval headquarters of their governments, which have powerful sending stations. One of the largest of these uses the Eiffel Tower in Paris as the support of its antenna, and sends out time signals to ships all over the Atlantic Ocean. Messages have been sent even as far as across the Atlantic Ocean by the Marconi stations at Wellfleet, Cape Cod, Massachusetts, and at Poldhu, England.

483. Wireless telephony. A wireless sending station ordinarily sends out wave trains at unvarying intervals, 1000 every second, because it is fed by an alternating current of unvarying frequency, say 500 cycles per second, and emits one wave train for every loop of the current. Since the telephone diaphragm of the receiving station moves once for each train received, its vibration is also at a uniform rate (1000 vibrations per second in the case just mentioned) and it emits a musical note of unvarying pitch. Recently sending stations have been devised that emit wave trains at varying intervals corresponding to the varying pitches and qualities of human speech. When such a succession of wave trains falls on an ordinary wireless receiving station the diaphragm in its telephone vibrates like the diaphragm of an ordinary telephone receiver, or like the diaphragm of a phonograph, and emits speech. Wireless telephone messages can be picked up by any one who has a properly tuned wireless telegraph set. Wireless telephony is already practicable over considerable distances, but is not yet (1913) a commercial success.

ELECTRICAL DISCHARGE THROUGH GASES

484. Sparking voltage. The voltage needed to make a spark jump between two knobs depends on several factors, such as the size of the knobs, the distance between them, and the atmospheric pressure. It takes less voltage to cause a spark to jump between two sharp points than between two round balls. For example, the **sparking voltage** for two sharp points 1 centimeter apart is about 7500 volts, and for two round balls 1 centimeter in diameter and 1 centimeter apart is about 27,000 volts. The sparking voltage between two sharp points varies so nearly as the distance that this is a method used to measure very high voltage.

To show the effect of atmospheric pressure we may connect a glass tube 2 or 3 feet long with an induction coil, as shown in figure 460. The tube is connected with a vacuum pump by a side tube. When the coil is first started, the discharge takes place between x and y, the terminals of the coil, which are only a few millimeters apart, but as the air is pumped out of the tube, the discharge goes through the long tube instead of across the short gap xy. This shows that the sparking voltage decreases when the pressure is diminished.

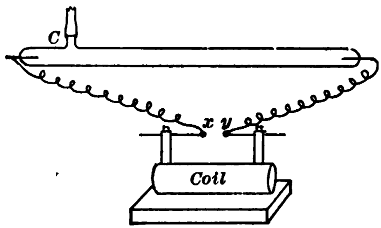


FIG. 460.—Discharge in partial vacuum.

485. Discharges in partial vacua. Reducing the atmospheric pressure between two points makes it easier for an electric discharge to pass, until a certain point in the exhaustion is reached. Then it begins to be more difficult. At the very highest degree of exhaustion yet attainable it is hardly possible to make a spark pass through a vacuum tube.

The changes in the appearance of such a tube as the exhaustion proceeds are very interesting. At first the discharge is along narrow flickering lines, but as the pressure

is lowered, the lines of the discharge widen out and fill the whole tube until it glows with a steady light. With still higher exhaustion, a soft, velvety glow covers the surface of the negative electrode or cathode, while most of the tube is filled with the so-called positive column which is luminous and stratified, and reaches to the anode. The so-called Geissler tubes (Fig. 461) are little tubes of this sort which are usually made in fantastic shapes and serve as pretty toys. The color of the light from a Geissler tube depends on the gas which is in the tube, and on the kind of glass used.

FIG. 461. — Geissler tube, made to study spectra of hydrogen.

486. Cathode rays. When the exhaustion of a tube is carried to a very high degree, so that the pressure is equal to about 0.0001 of a millimeter of mercury, the positive glow is very faint and the dark space around the cathode is pervaded by a discharge. An invisible radiation streams out nearly at right angles to the cathode surface, no matter where the anode is located in the tube. This radiation from the cathode is called **cathode rays** and shows itself in several ways: *first* by a yellowish green fluorescence wherever it strikes the glass of the tube; *second*, by the fact that it can be brought to a focus where it produces intense heat; and *third*, by the sharply defined shadows which a metal interposed in its path produces in the fluorescence on the end of the tube.

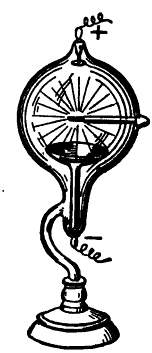


FIG. 462. — Heating effect of cathode rays.

A Crookes' tube, arranged as in figure 462, shows the heating effect of the cathode rays. When an induction coil sends a discharge through the tube from top to bottom, the cathode rays are focused on a piece of platinum which becomes red hot.

Another Crookes' tube, arranged as in figure 463, shows that a shadow is formed on the end of the tube by an aluminum cross.

487. Bending of cathode rays. A Crookes' tube, made as in figure 464, sends a narrow band of cathode rays through the slit *s* in the aluminum screen *mn* against a fluorescent screen *f* slightly inclined to them. When a strong magnet *M* is held near the side of this tube, it is found that the stream of cathode rays is deflected in the direction which would be expected if they were a stream of negatively charged particles. From this and other experiments we believe that *cathode rays are negatively charged particles projected at very high velocity from the cathode.*

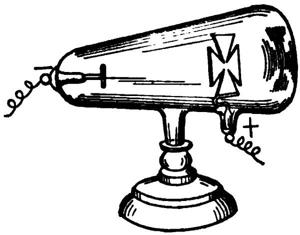


FIG. 463. — Shadow formed by cathode rays.

J. J. Thomson, the English physicist, has estimated from various experiments on cathode rays that the negatively charged particles, which he calls **electrons**, have each a mass about sixteen hundred times smaller than that of a hydrogen atom, and move with a velocity of from one tenth to one third that of light. It is supposed that each particle carries a negative charge of electricity equal to that of the hydrogen atom in electrolysis.

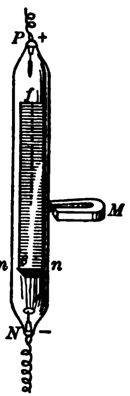


FIG. 464. — Bending of cathode rays by a magnet.

488. Roentgen rays. In 1895, while experimenting with a vacuum tube, Roentgen discovered another kind of rays which he called X-rays. When cathode rays strike against a platinum target, as shown in figure 465, Roentgen rays are sent off from this target. They affect a photographic plate somewhat as sunlight does; but, like cathode rays, they will penetrate

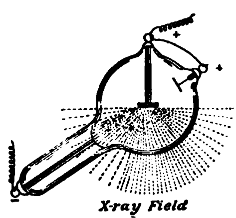


FIG. 465. — Roentgen ray tube.

many substances opaque to ordinary light, such as wood, pasteboard, and the human body. That they are not the same as cathode rays is shown by the fact that they are not deflected by a magnet.

When a photographic plate, inclosed in the usual plate-holder with sides of hard rubber or pasteboard, is exposed, with a hand held over it, to Roentgen rays, a shadow picture like that seen on the fluorescent screen is formed.

We may demonstrate the action of Roentgen rays by operating an X-tube with an induction coil, and holding a fluorescent screen in front of the bulb. If the room is dark and the hand is interposed between the tube and the screen, the flesh, which is easily penetrated by the rays, will be seen faintly outlined, while the bones will cast a strong shadow.

Figure 466 (opposite page 478) is from a photograph taken by means of X-rays, and shows how valuable they are to doctors.

Roentgen rays are produced at, and sent forth from, any solid body upon which cathode rays fall. They are now known to be *ether waves*, just like light waves and wireless telegraph *waves*, but of very short wave length.

489. Radioactivity. Near the end of the nineteenth century, scientists discovered that something which resembles Roentgen rays is radiated from certain rare minerals, such as uranium, pitch-blende, and thorium. It affects a photographic plate through an envelope of black paper. It has also the power of discharging electrified bodies, and so by using a very sensitive electroscope it is possible to detect and measure the intensity of this radiation. This new phenomenon is called **radioactivity**, and a new element, which is remarkably radioactive, has been discovered and called **radium**. In this interesting and novel field of research many scientists are now seeking to learn the answer to the great questions "What is electricity?" and "What is inside the atoms of substances?"

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